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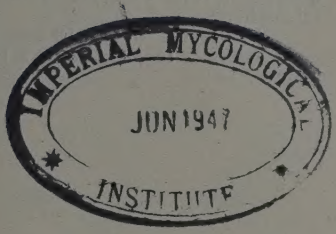

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MATHÉMATIQUES

GÉNÉRALISATION D'UN THÉORÈME GÉOMÉTRIQUE DE M. SERGE BERNSTEIN

Par G. ADELSON-WELSKY

(Présenté par S. N. Bernstein, de l'Académie, le 2. VI. 1945)

M. Serge Bernstein [1] a démontré le théorème suivant:

Si $f(x, y)$ est une fonction dont les dérivées partielles des deux premiers ordres sont continues et si la courbure totale de la surface $z = f(x, y)$ est non positive et non identiquement nulle, la surface ne peut pas rester toujours contenue entre deux plans fixes $z = \pm h$.

Dans la suite M. Bernstein a généralisé son théorème en démontrant [2] qu'il subsiste, si au lieu d'exiger que $z = f(x, y)$ soit bornée, on suppose qu'elle croît assez lentement, à savoir, si

$$\lim_{\sqrt{x^2+y^2}\to\infty} \frac{|f(xy)|}{\sqrt{x^2+y^2}} = 0 \qquad (1)$$

Nous allons démontrer un théorème analogue sans exiger que $f(x, y)$ soit dérivable.

Définition. Une surface σ définie par l'équation $z = f(x, y)$ possède une courbure généralisée non positive, si pour chaque plan γ les ensembles ouverts G^+, des points de γ situés au-dessus de la surface σ, et G^-, des points de γ situés au-dessous de σ, ne possèdent pas de composantes bornées et connexes.

Le théorème suivant a lieu.

Si une fonction $y = f(x, y)$, qui définit une surface σ à courbure généralisée non positive, croît plus lentement qu'une fonction linéaire quelconque, c'est-à-dire si le rapport $|f(x, y)| / \sqrt{x^2+y^2}$ tend vers zéro quand $\sqrt{x^2+y^2}$ croît, la surface σ est une surface cylindrique dont les génératrices sont parallèles au plan $[x, y]$.

Lemme 1. Si sur une surface σ à courbure généralisée non positive il existe trois points a, b et c, situés dans un même plan vertical, mais pas sur une même droite, par chaque droite l, située dans ce plan vertical et telle que le point médian b est d'un côté et les points extrêmes a et c de l'autre côté de cette droite, il passe un plan pour lequel l'un des domaines G^+ ou bien G^- est non connexe, et les projections verticales des points a et c sont situées dans deux composantes différentes de G^+ (resp. G^-).

Soit σ une surface à courbure généralisée non positive définie par une fonction continue $z = f(x, y)$, qui croît plus lentement qu'une fonction linéaire quelconque. Posons

$$v(\rho) = \max_{\sqrt{x^2+y^2}\leq\rho} |f(x, y)| \qquad (2)$$

Alors

$$\frac{v(\rho)}{\rho} \to 0 \qquad \text{pour } \rho \to \infty \qquad (3)$$

Menons un plan γ non parallèle au plan $[x, y]$ et tel que l'un de deux ensembles ouverts G^+ ou G^- soit non connexe. En transférant parallèlement et en effectuant une rotation des axes de coordonnées nous pouvons obtenir que l'équation de ce plan devienne $z = py$.

Lemme 2. Il existe pour chaque $\delta\ (p > \delta > 0)$ un nombre R tel que le domaine du plan γ, défini par les inégalités

$$y > \delta |x| / \sqrt{p^2 - \delta^2} \quad \text{et} \quad x^2 + y^2 > R^2,$$

soit entièrement situé dans le domaine G^+, et le domaine défini par les inégalités $y < -\delta |x| / \sqrt{p^2 - \delta^2}$ et $x^2 + y^2 > R^2$, soit entièrement situé dans le domaine G^-.

Comme l'un des domaines G^+ et G^- est non connexe, l'une des composantes de l'un de ces domaines est située dans la bande entre les domaines définis dans l'énoncé du lemme 2. Soit Ω cette composante connexe et supposons, pour fixer les idées, $\Omega \subset G^-$. Sur chaque droite $x = \text{const}$ l'ensemble des valeurs de y pour lesquelles $(x, y) \subset \Omega$ est borné, et la fonction $z' = f(x, y) - py$ atteint sa valeur maximum, car, bien que cet ensemble n'est pas fermé, mais en ses points frontières on a $f(x, y) - py = 0$. Posons

$$u(x) = \max_{x = \text{const};\ (x, y) \in \Omega} (f(x, y) - py) \tag{4}$$

Lemme 3. Dans les hypothèses faites la fonction $u(x)$ est une constante.

Il résulte immédiatement du lemme 3 que le domaine Ω est infini dans les deux sens.

Lemme 4. La surface σ contient une droite entière $y = y_1$, $z = u + py_1$. On a ici $u = u(x) = \text{const}$.

Désignons par S l'ensemble des droites situées sur la surface σ et parallèles à la droite obtenue $y = y_1$, $z = u + py_1$. L'ensemble S est fermé.

Théorème. *Dans les hypothèses faites concernant la surface σ définie par la fonction $z = f(x, y)$, cette surface est une surface cylindrique dont les génératrices sont parallèles à l'axe x.*

Si un ensemble fermé S ne coïncide pas avec tout le plan, le complémentaire à cet ensemble contient une bande ou bien un demi-plan. On démontre que dans cette bande ou dans ce demi-plan on peut mener encore une droite située sur la surface σ et parallèle à l'axe x. Cela contredit à ce que cette droite ou ce demi-plan est contenu dans le complémentaire à l'ensemble S.

Université d'État Lomonosov.
Moscou.

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MATHÉMATIQUES

QUELQUES APPLICATIONS DE L'INTÉGRALE DE VALLÉE-POUSSIN A LA THÉORIE DES SÉRIES DE FOURIER

Par I. NATANSON

(Présenté par S. L. Soboleff, de l'Académie, le 10. I. 1945)

§ 1. Soit

$$V_n(x) = \frac{(2n)!!}{(2n-1)!!}\,\frac{1}{2\pi}\int_{-\pi}^{\pi} f(t)\cos^{2n}\frac{t-x}{2}\,dt \tag{1}$$

c'est l'intégrale de Vallée-Poussin.

On sait que

$$\cos^{2n}\frac{u}{2} = \frac{(2n-1)!!}{(2n)!!}\left[1 + 2\sum_{k=1}^{n}\frac{(n!)^2}{(n-k)!\,(n+k)!}\cos ku\right]$$

Alors

$$V_n(x) = \frac{q_0^{(n)}a_0}{2} + \sum_{k=1}^{n} q_k^{(n)}(a_k\cos kx + b_k\sin kx) \tag{2}$$

où

$$q_k^{(n)} = \frac{(n!)^2}{(n-k)!\,(n+k)!} \qquad (k = 0,\ 1,\ \ldots,\ n) \tag{3}$$

et

$$\frac{a_0}{2} + \sum_{k=1}^{\infty}(a_k\cos kx + b_k\sin kx) \tag{4}$$

est la série de Fourier de $f(x)$.

Alors l'intégrale (1) est le résultat de l'application à la série (4) de la méthode de sommation au moyen des multiplicateurs (3). Nous allons nommer cette méthode de sommation «méthode V».

Théorème 1. *La méthode V est plus forte que la méthode* $(C, 2)$:

En effet, une méthode de sommation des séries au moyen des multiplicateurs $\{q_k^{(n)}\}$ $(k = 0,\ 1,\ \ldots,\ n)$ n'est pas plus faible que $(C, 2)$, si

$$\lim_{n\to\infty} q_k^{(n)} = 1 \tag{5}$$

et

$$\sum_{k=1}^{n} |\,q_k^{(n)} - 3q_{k+1}^{(n)} + 3q_{k+2}^{(n)} - q_{k+3}^{(n)}\,|\,(k+1)(k+2) < C \tag{6}$$

Il est évident que la condition (5) est remplie. On peut vérifier la condition (6) en s'appuyant sur l'identité

$$\Delta_3 q_k^{(n)} = q_k^{(n)} - 3q_{k+1}^{(n)} + 3q_{k+2}^{(n)} - q_{k+3}^{(n)} = \\ = 2q_k^{(n)} \frac{(2k+3)(2k^2+6k+1-3n)}{(n+k+1)(n+k+2)(n+k+3)}$$

d'où nous pouvons conclure que $\Delta_3 q_k^{(n)} \geqslant 0$ pour

$$k \geqslant \frac{\sqrt{6n+7}-3}{2}$$

et $\Delta_3 q_k^{(n)} < 0$ pour les autres valeurs de k.

Ainsi la méthode V n'est pas plus faible que la méthode $(C, 2)$.

D'autre part, si nous posons

$$\sum_{k=0}^{n} q_k^{(n)} a_k = V_n$$

on peut démontrer que pour $n > 0$

$$a_n = (-1)^n 2n \sum_{k=0}^{n} (-1)^k \frac{(n+k-1)!}{(k!)^2 (n-k)!} V_k$$

Cela étant, il est aisé de construire une série sommable par la méthode V et non-sommable par la méthode d'Abel-Poisson.

Si nous appliquons ce théorème aux séries de Fourier, nous obtiendrons le

Théorème 2. *Soit $f(x)$ une fonction intégrale L. Si en un point x la série de Fourier de cette fonction est sommable par la méthode $(C, 2)$ (avec la somme généralisée $f(x)$), alors en ce point on a*

$$\lim_{n\to\infty} V_n(x) = f(x)$$

§ 2. Soit $f(z)$ une fonction périodique (de période 2π) de la variable complexe z. On sait que, si cette fonction est holomorphe dans une bande du plan z

$$b < \mathrm{I}(z) < B \qquad (b < 0,\ B > 0) \tag{7}$$

alors elle est développable en série de Fourier. D'où nous obtenons le

Théorème 3. *Si $f(z)$ est holomorphe dans la bande (7) et $f(z+2\pi) = f(z)$, alors dans tous les points z de cette bande la relation*

$$\lim_{n\to\infty} V_n(z) = f(z)$$

a lieu. En particulier, cette relation a lieu en tous les points du plan z, si $f(z)$ est une fonction entière.

§ 3. Soit donnée une série trigonométrique (4). On dit que cette série appartient à la classe C, M, L_p, L, si elle est la série de Fourier d'une fonction appartenant à la classe correspondante. Si la série (4) est la série de Fourier-Stjelties, on dit qu'elle appartient à la classe S.

Au moyen de l'intégrale de Vallée-Poussin nous pouvons donner quelques critères nouveaux pour déterminer la classe d'une série trigonométrique.

Théorème 4. *Soit donnée une série (4). Posons*

$$V_n(x) = \frac{q_0^{(n)} a_0}{2} + \sum_{k=1}^{n} q_k^{(n)} (a_k \cos kx + b_k \sin kx)$$

Pour que la série (4) *appartienne à la classe* K, *il faut et il suffit que*

1) *Pour* $K=C$, *les fonctions* $V_n(x)$ *soient également continues.*

2) *Pour* $K=M$, *les fonctions* $V_n(x)$ *soient uniformément bornées.*

3) *Pour* $K=L_p\ (p>1)$ *les intégrales*

$$\int_{-\pi}^{\pi} |V_n(x)|^p\, dx$$

soient uniformément bornées.

4) *Pour* $K=L$, *la relation*

$$\lim_{n,\, m\to\infty} \int_{-\pi}^{\pi} |V_n(x)-V_m(x)|\, dx = 0$$

ait lieu.

5) *Pour* $K=S$, *les intégrales*

$$\int_{-\pi}^{\pi} |V_n(x)|\, dx$$

soient uniformément bornées.

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FLUID DYNAMICS

ON THE POSSIBILITIES OF A THEORETICAL STUDY OF CAVITATION AS A MOTION OF A SPECIAL KIND OF COMPRESSIBLE FLUID

By L. A. EPSTEIN

(Communicated by N. E. Kochin, Member of the Academy, 16. XII. 1944)

By cavitation is described a complex of phenomena arising in a dropping fluid when local static pressures are approaching zero (vanishing). Changes in the flow take place owing to evolution of vapour and gases dissolved in the liquid. Cavitation sharply changes the force exerted by the liquid on the body past which it is flowing, and brings on a series of specific effects such as failure of materials, the appearance of noises, vibrations, *etc.*

Heretofore cavitation has been computed and the problem of this phenomenon discussed only for jet (or sheet) flows, which correspond to the smallest possible cavitation numbers

$$\lambda = \frac{2(p_0 - p_d)}{\rho v_0^2} \tag{1}$$

where v_0 and p_0 are the velocity and pressure far away from the body, p_d is the pressure of water vapour at a given temperature, and ρ is the density of the fluid.

The action of the flow on the body in the stage of cavitation is studied by purely experimental methods in variable pressure water tunnels. It is taken for granted that the flow past an experimental model is true to nature if the value of λ is kept constant. Cavitation tests are usually conducted at velocities and pressures that are lower than those taking place under natural conditions. From my own observations of cavitations flows and those reported by other authors [1-3] it seems certain that during the succession of changing states that begin with the appearance of cavitation and end into violent boiling and complete separation, the flow is unhomogeneous in composition presenting a liquid with a great many small bubbles of gas or vapour.

Visually the region of cavitation is perceived as a foggy cloud that appears in a flowing transparent liquid and expands as the cavitation number grows smaller. Its visibility is due to the refraction and reflection of light by the bubbles.

The photographs in Fig. 1 show the development of the cavitation in a narrow tube with a neck.

We shall proceed on the assumption that this unhomogeneous flow is a stream of a hypothetical homogeneous fluid.

On physical considerations the density ρ_1 of this fluid must be definitely related not merely to the pressure but also to the velocity and other elements of the flow. If the negligibly small mass of the liberated gases and vapour be disregerded, we shall have

$$\rho_1 = \rho_0 V$$

where V is the volume ratio between the dropping and the hypothetical fluid.

At a given temperature the quantity ρ_1 depends in the first place on pressure p_1. If p_1 is not too small, there is no evaporation within the fluid, or even any appreciable evolution of the dissolved gases, so that $V=1$ and $\rho_1=\rho_0$.

With p_1 dropping a portion of the volume will be occupied by gases evolving according to Henry Dalton's law, and then, also, by vapour. Accordingly, V and ρ_1 will be diminished.

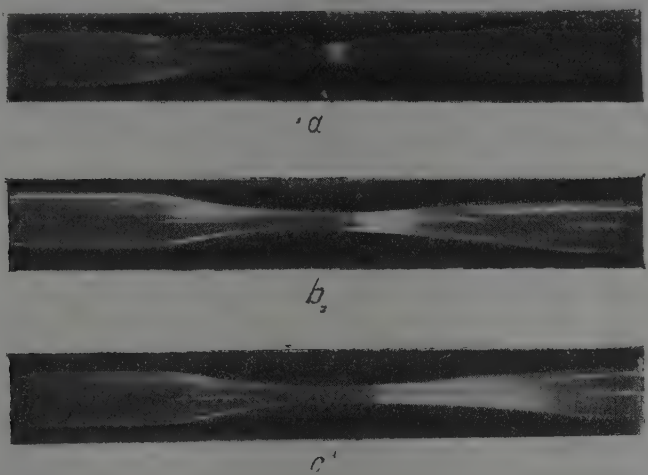

Fig. 1

The volume of the vapour and of the gases disengaged from solution depends also upon the time t during which the particles of the fluid traverse the region of reduced pressure, and upon the rate of evaporation * (separation from solution) u. Hence

$$\rho_1=\rho_1(p,\ t,\ u,\ \ldots) \tag{2}$$

It will be obvious that in constructing a model of cavitation phenomona one has to secure the invariability not only of the cavitation number λ, but also of the dimensionless quantity

$$\sigma=\frac{ul}{v} \tag{3}$$

where l is a characteristic linear size. It can be proved that for large values of $\sigma\gg\sigma_{cr}$ this quantity does not play any essential rôle, and cavitation in this region is modelled faithfully when only λ is constant. In practice we often have $\sigma\ \sigma_{cr}$, which accounts for the fact that similarity betwen the model and nature is seldom broken if λ =const. The deviations may generally be attributed to the inaccuracy of the experiment or some incidental factors.

When $\sigma\gg\sigma_{cr}$ the quantity ρ can be regarded as variable only with pressure. On Henry Dalton's law and taking also into consideration the character of the thermodynamic process, the form of the functions $\rho\ (p_1)$ may be established as follows

$$\rho_1=\rho_0\frac{1}{1+\delta_0\left(\frac{p_0-p_d}{p_1-p_d}-1\right)}\approx\frac{1}{1+\delta_0\frac{p_0-p_d}{p_1-p_d}} \tag{4}$$

where δ_0 is the specific volume of the gases dissolved in the fluid (for the water $\delta_0\approx 2$ per cent).

* The rate of evaporation (separation from solution) u can be understood to mean the mass of vapour (or gas) evolved in unit time from a unit mass of the liquid. This quantity depends strongly on the content of dust, gas bubbles and other alien particles in the liquid, whereby the processes of gas evolution and evaporation are facilitated.

It will be apparent that the formation and development of cavitation can be studied by the methods of gas dynamics. Some of the results obtained in that domain, which has been rather thoroughly investigated, can be applied directly to the problem of cavitation, others are to be modified in accordance with the different form of relation between density and pressure.

To take the simplest examples of cavitation research one may consider a unidimensional steady motion of the hypothetical fluid in a tube and past a thin wing.

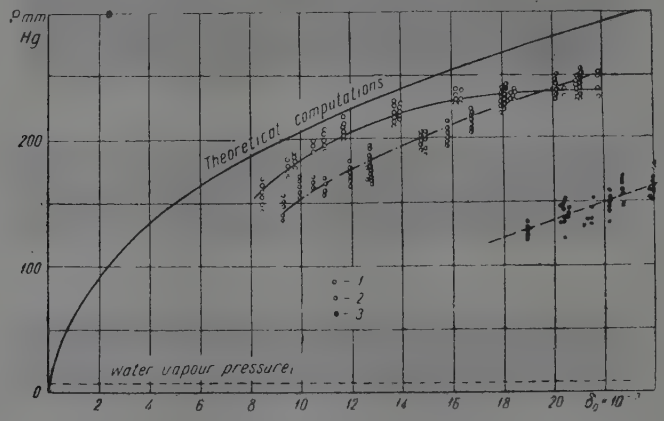

Fig. 2. Experiments by Numachi: *1*, sea water; *2*, salt water; *3*, distilled water

In place of Bernoulli's equation and the equation of continuity which ordinarily connect pressure velocity and cross-section area, we now have, taking into account the compressibility of the hypothetical fluid,

$$v = \sqrt{1 + \lambda - \lambda p - \delta_0 \lambda \ln p}, \qquad S = \frac{1 + \delta_0 / p}{v} \tag{5}$$

where v, p, S are the ratios of the local values of velocity v_1, excessive pressure $p_1 - p_d$ and area S_1 to the respective values at infinity.

It follows from expression (5) that $S(p)$ has a minimum and $v(p)$ increases indefinitely as p tends to zero. When $S \geqslant S_{\min}$, the velocity v is equal to, or greater than, the local velocity of sound a

$$a = \sqrt{\frac{p_0}{\rho \delta_0} (p + \delta)}$$

The region of cavitation flow is the one in which the hypothetical fluid is moving with supersonic velocity. The end of the cavitation region corresponds to the sudden increase in density. At the place where the cavitation is initiated, the pressure is many times as great as the pressure of water vapour

$$p_1 \approx p_0 \sqrt{2\delta_0} \left(1 - \sqrt{2\delta_0}\right) + p_d \tag{6}$$

In the narrow tube with a given constriction the cavitation arises when

$$\lambda_{cr} = \frac{1 - S^2}{S^2} \frac{1}{\left(1 - \sqrt{2\delta_0}\right)^2} \tag{7}$$

These theoretical computations can be compared with the results of Numachi's experiments on the appearance of cavitation in narrow glass tubes ([3]). He used distilled water and ordinary sea water to determine the relation between the pressure at the sitewhere cavitation appears and the amount of air dissolved in water.

Fig. 2 shows that the results he obtained are in good qualitative agreement with theoretical calculations*. Quantitatively the calculation agrees with the experiment only for sea water. In distilled water the pressures turn out to be much lower. But far from contradicting the theory this evidence only goes to confirm the above considerations. In sea water which, indeed, contains foreign matter, microorganisms, *etc.* u and σ are large enough for the evolution of air from solution to obey Henry Dalton's laws. But in distilled water the rate of evolution is insufficient, and only a small portion of the dissolved air is set free. It is now possible to account for the divergence between the data of the above-mentioned experiments and for their deviation from Hunsaker's observations, who ([4]) obtained still lower pressures because his experiments were run at considerably smaller values of σ.

In the adjoined table are given the computed ratios of the dissolved gas evolution (evolution according to Henry Dalton's law is put at 100 per cent).

Numachi's experiments			Hunsaker's experiment
Sea water	Salt water	Distilled water	Distilled water
78	75	20	1—2

It may be inferred from this table that in studying cavitation in real water one can in many cases use the assumptions placed at the basis of the theory of a hypothetic compressible fluid, developed for the case of $\sigma \gg \sigma_{cr}$. On the other hand, for great velocities of motion, that is for $\sigma \leqslant \sigma_{cr}$, the theory must be worked out to greater complexity by determining the form of the function ρ. In this case reference to λ only will not suffice to obtain a correct model of the cavitation phenomenon.

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* In Fig. 2 Numachi's experiments are compared with the pressure which would have been obtained for the theoretically established discharge if calculated for an incompressible fluid according to Bernoulli's equiation and the equation of continuity. This is the way in which Numachi treated his data.

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THEORY OF ELASTICITY

ON RAYLEIGH BOUNDARY WAVES

By V. GOGOLADZE

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In the present note we investigate the spectrum of free vibrations of two different solid elastic media meeting along a plane, and establish a necessary and sufficient condition for the existence of Rayleigh's boundary wave.

Using the notations of the preceding note (1), we write the potentials of elastic waves in the form

$$\left.\begin{aligned} \varphi_1 &= G_1 f_j\left(pt+\alpha x \pm iy\sqrt{\alpha^2+\frac{p^2}{a_1^2}}\right)\\ \psi_1 &= G_2 f_j\left(pt+\alpha x \pm iy\sqrt{\alpha^2+\frac{p^2}{b_1^2}}\right)\end{aligned}\right\} y>0$$

$$\left.\begin{aligned} \varphi_2 &= G_3 f_j\left(pt+\alpha x \mp iy\sqrt{\alpha^2-\frac{p^2}{a_2^2}}\right)\\ \psi_2 &= G_4 f_j\left(pt+\alpha x \mp iy\sqrt{\alpha^2-\frac{p^2}{b_2^2}}\right)\end{aligned}\right\} y<0 \qquad (1)$$

$$j=1,2$$

where f_1 is a regular function in the upper complex half-plane, f_2 is regular in the lower complex half-plane, $|f_j|<M$ $(M>0)$ and $f_1(z)=f_2(z)$. The upper sign before the radicals in (1) is valid for $j=1$, and the lower sign, for $j=2$. The common free vibrations satisfiyng the conditions of stresses and displacements in the plane $y=0$ lead us to Rayleigh's equation

$$\begin{aligned} R(\vartheta)=&\left[\vartheta^2\left(\frac{1}{b_1^2}-\frac{2}{b_2^2}\right)-2(1-\sigma)\right]^2-\left[\sqrt{1-\frac{\vartheta^2}{a_1^2}}\sqrt{1-\frac{\vartheta^2}{b_2^2}}+\right.\\ &\left.+\sqrt{1-\frac{\vartheta^2}{a_2^2}}\sqrt{1-\frac{\vartheta^2}{b_1^2}}\right]^2\frac{\vartheta^4}{b_1^2b_2^2}-\sqrt{1-\frac{\vartheta^2}{a_1^2}}\sqrt{1-\frac{\vartheta^2}{b_1^2}}\left(\frac{2\vartheta^2}{b_2^2}+2(1-\sigma)\right)^2-\\ &-\sqrt{1-\frac{\vartheta^2}{a_2^2}}\sqrt{1-\frac{\vartheta^2}{b_2^2}}\left(\frac{\vartheta^2}{b_1^2}-2(1-\sigma)\right)^2+\\ &+4(1-\sigma)^2\sqrt{1-\frac{\vartheta^2}{a_1^2}}\sqrt{1-\frac{\vartheta^2}{a_2^2}}\sqrt{1-\frac{\vartheta^2}{b_1^2}}\sqrt{1-\frac{\vartheta^2}{b_2^2}}=0 \end{aligned} \qquad (2)$$

where $\vartheta=p/|\alpha|$ is the phase velocity of the vibration.

We investigate the roots of Rayleigh's equation in the whole plane of the variable ϑ by the method briefly indicated in the note (2). The velocities $b_j<a_j$ and $\sigma=\mu_2/\mu_1$ may have arbitrary values. Let, for instance, the velocity distribution be

$$b_2<b_1<a_1<a_2 \qquad (3)$$

Cutting the plane ϑ along the segment $(-a_2, a_2)$, we take for the first sheet of the Riemanns surface the one in which for $\vartheta = i\eta$ $(\eta > 0)$ the radicals entering into $R(\vartheta)$ assume positive values.

It may be shown that on the first sheet of Riemann's surface in the neighbourhood of $\vartheta = 0$ we have $R(\vartheta) < 0$ and $R(\vartheta) \sim \vartheta^4$ for real values of ϑ.

In order to find the roots of $R(\vartheta)$ on the first sheet of Riemann's surface, we investigate the variation of its argument along the closed contour (C) consisting of a segment of the real axis and a semi-circle of a large radius r in the lower complex half-plane. For $\vartheta > a_2$ we have

$$R(\vartheta) = \left[\vartheta^2\left(\frac{1}{b_1^2} - \frac{\sigma}{b_2^2}\right) - 2(1-\sigma)\right]^2 + \left[\sqrt{\frac{\vartheta^2}{a_1^2} - 1}\sqrt{\frac{\vartheta^2}{b_2^2} - 1} + \right.$$
$$\left. + \sqrt{\frac{\vartheta^2}{a_2^2} - 1}\sqrt{\frac{\vartheta^2}{b_1^2} - 1}\right]\frac{\sigma\vartheta^4}{b_1^2 b_2^2} + \sqrt{\frac{\vartheta^2}{a_1^2} - 1}\sqrt{\frac{\vartheta^2}{b_1^2} - 1}\left(\frac{\sigma\vartheta^2}{b_2^2} + 2(1-\sigma)\right)^2 +$$
$$+ \sqrt{\frac{\vartheta^2}{a_2^2} - 1}\sqrt{\frac{\vartheta^2}{b_2^2} - 1}\left(\frac{\vartheta^2}{b_1^2} - 2(1-\sigma)\right)^2 +$$
$$+ 4(1-\sigma)^2\sqrt{\frac{\vartheta^2}{a_1^2} - 1}\sqrt{\frac{\vartheta^2}{b_1^2} - 1} - \sqrt{\frac{\vartheta^2}{a_2^2} - 1}\sqrt{\frac{\vartheta^2}{b_2^2} - 1} \qquad (4)$$

and consequently in the interval

$$a_2 < \vartheta < \infty \quad \arg R(\vartheta) = 0$$

In the intervals (a_1, a_2) and (b_1, a_1) $R(\vartheta)$ is a complex quantity, whose argument varies in the lower complex half-plane. When the variable ϑ describes a closed contour around the point of ramification $\vartheta = b_1$ and varies in the interval (b_2, b_1), the function of Rayleigh will be expressed in the following form

$$R(\vartheta) = P(\vartheta) - iQ(\vartheta)\sqrt{\frac{\vartheta^2}{b_2^2} - 1} \qquad (5)$$

where

$$P(\vartheta) = \left[\vartheta^2\left(\frac{1}{b_1^2} - \frac{\sigma}{b_2^2}\right) - 2(1-\sigma)\right]^2 - \frac{\sigma\vartheta^4}{b_1^2 b_2^2}\sqrt{1 - \frac{\vartheta^2}{b_1^2}}\sqrt{1 - \frac{\vartheta^2}{a_2^2}} -$$
$$- \sqrt{1 - \frac{\vartheta^2}{a_1^2}}\sqrt{1 - \frac{\vartheta^2}{b_1^2}}\left(\frac{\sigma\vartheta^2}{b_2^2} + 2(1-\sigma)\right)^2 \qquad (6)$$

$$Q(\vartheta) = \left[-4(1-\sigma)^2\sqrt{1 - \frac{\vartheta^2}{a_1^2}}\sqrt{1 - \frac{\vartheta^2}{b_1^2}}\sqrt{1 - \frac{\vartheta^2}{a_2^2}} + \right.$$
$$\left. + \sqrt{1 - \frac{\vartheta^2}{a_2^2}}\left(\frac{\vartheta^2}{b_1^2} - 2(1-\sigma)\right)^2 + \sqrt{1 - \frac{\vartheta^2}{a_1^2}}\frac{\sigma\vartheta^4}{b_1^2 b_2^2}\right] \qquad (7)$$

Let $\sigma > 1$, then $Q(\vartheta) > 0$ in the interval (b_2, b_1). But if $\sigma < 1$, then, ϑ varying in the interval (b_2, b_1), a change of the sign of the functions $P(\vartheta)$ and $Q(\vartheta)$ is possible. To investigate $\arg R(\vartheta)$ in this interval, we study the properties of the functions (6) and (7) in the plane determined by means of the elastic constants

$$\xi = 2(1-\sigma) = 2\left(1 - \frac{\mu_1}{\mu_2}\right), \quad \eta = \delta = \frac{\rho_2}{\rho_1} \qquad (8)$$

Between σ and δ there exists the obvious relation $\delta = \sigma b_1^2 / b_2^2$ which in the plane (ξ, η) defines the straight line

$$\eta = \frac{b_1^2}{b_2^2}\left(1 - \frac{\xi}{2}\right), \quad \frac{b_1^2}{b_2^2} > 1 \qquad (9)$$

Substitute into (6) and (7) $\vartheta^2/b_1^2=\zeta^2$ and consider the curves

$$P(\xi,\eta,\zeta)=[\zeta^2\eta+\xi]^2-2\zeta^2[\zeta^2\eta+\xi]+\zeta^4-\sqrt{1-\zeta^2}\sqrt{1-\frac{b_2^2}{a_2^2}\zeta^2}\,\zeta^4\eta-$$
$$-\sqrt{1-\frac{b_1^2}{a_1^2}\zeta^2}\sqrt{1-\zeta^2}\,(\zeta^2\eta+\xi)^2=0 \qquad (10)$$

$$Q(\xi,\eta,\zeta)=-\xi^2\sqrt{(1-\zeta^2)\left(1-\frac{b_1^2}{a_1^2}\zeta^2\right)\left(1-\frac{b_1^2}{a_2^2}\zeta^2\right)}+$$
$$+\sqrt{1-\frac{b_1^2}{a_1^2}\zeta^2}\,\zeta^2\zeta^4\eta-\sqrt{1-\frac{b_1^2}{a_2^2}\zeta^2}\,\zeta^2(\zeta^2-\xi)^2=0 \qquad (11)$$

where $b_2^2/b_1^2<\zeta^2<1$.

These curves have common points of intersection $(\xi_1, 0)$, $(\xi_2, 0)$ and

$$\xi_3=\xi_4,\quad \eta_3=\eta_4>0 \qquad (12)$$

On ground of (12) the parabolae $P=0$ and $Q=0$ intersect on the axis of abscissae at the points $(\xi_1, 0)$, $(\xi_2, 0)$ and touch each other in the domain $\eta<0$ at the point $\xi_3=\xi_4$, $\eta_3=\eta_4$. The axis of the parabola $Q=0$ is parallel to the axis of η and it opens towards negative η. The parabola $P=0$ intersects the axis $\xi=0$ at real points situated in $\eta>0$ and $\eta'_{\xi=\xi_1}<0$. In the domain $\eta=\delta=\rho_2/\rho_1>0$ the parabola $Q=0$ lies inside the parabola $P=0$. The point $(0,0)$ is situated outside the parabolae (10) and (11), and $P(0,0)>0$; thus, inside the parabola (10) $P(\xi,\eta,\zeta)<0$. Outside the parabola (11) $Q(\xi,\eta,\zeta)>0$.

Thus, for $P(\xi,\eta,\zeta)=0$ we have $Q(\xi,\eta,\zeta)>0$. On ground of the relations established between the curves (10) and (11) we may now state the following theorem:

When ϑ varies in the interval (b_2, b_1) from the end-point $\vartheta=b_1$ to b_2, the function $P(\vartheta^2)$ changes its sign earlier than the function $Q(\vartheta)$.

An elementary study of the properties of the functions $P(\vartheta)$ and $Q(\vartheta)$ on the plane ϑ leads us to the following result: these functions may have in the interval (b_2, b_1) only one root each.

Now we may investigate the roots of Rayleigh's function $R(\vartheta)$ in the whole plane ϑ.

Assume that $R(b_2)>0$. This means that the function $P(\vartheta)$ does not change its sign in the interval (b_2, b_1) and hence the increment of $\arg R(\vartheta)$ in the intervals (b_2, ∞) and $(-\infty, -b_2)$ is equal to zero. Since $R(b_2)>0$ and in virtue of the behaviour of $R(\vartheta)$ in the neighbourhood of $\vartheta=0$, $R(\vartheta)$ must have at least one root in the interval $(0, b_2)$ and one symmetric to it in the interval $(-b_2, 0)$. The variation of $\arg R(\vartheta)$ on a contour enclosing $\vartheta=0$ is equal to -4π in view of the property of $R(\vartheta)$. Hence the variation of $\arg R(\vartheta)$ along the real axis from $+\infty$ to $-\infty$ is equal to -6π.

On a semi-circle of a large radius in the lower half-plane ϑ $R(\vartheta)$ is of the order ϑ^6. Thus, the complete variation of $\arg R(\vartheta)$ along the closed contour (C) is equal to zero. The assumption of the existence of complex roots, or more than one real root in the interval $(0, b_2)$ leads us to a contradiction, because then the increment of $\arg R(\vartheta)$ along the closed contour (C) should be negative, which is impossible, $R(\vartheta)$ having no poles. Let now $R(b_2)<0$. Then $P(\vartheta)$ changes its sign in the interval (b_2, b_1) once, and for $P(\vartheta)=0$, $Q(\vartheta)>0$. Hence the increment of $\arg R(\vartheta)$ in this interval is equal to $-\pi$.

Since at the end-points of the interval $(0, b_2)$ $R(\vartheta)$ is negative, there are either no roots of $R(\vartheta)$ in the interval $(0, b_2)$, or there will be not less than two roots. If we have n roots, then the increment of $\arg R(\vartheta)$ along the closed contour (C) will be equal to $2\pi(1-n)<0$ for $n\neq 0$.

The assumption of the existence of a complex root also leads to a negative sign of the increment of arg $R(\vartheta)$, which implies a contradiction. Hence, if $R(b_2) < 0$, the function of Rayleigh $R(\vartheta)$ has no roots on the first sheet of Riemann's surface.

The velocity distributions different from (3) may be investigated by the same method. Thus, for the existence of a point spectrum of free vibrations (Rayleigh waves) it is necessary and sufficient that the following conditions should be satisfied

$$R(b_2) > 0 \text{ for } b_2 < b_1, \text{ or } R(b_1) > 0 \text{ for } b_1 < b_2 \tag{13}$$

These conditions determine a class of domains of elastic media, for which Rayleigh waves arise. The points of the plane (ξ, η) which are situated outside the parabola $P(\xi, \eta) = 0$ and on the straight line (9) determine the Rayleigh waves. For points situated inside $P = 0$ no Rayleigh waves arise.

If, for instance, $R(b_2) > 0$ $(b_2 < b_1)$, then there exists a root of Rayleigh equation

$$R(\pm \varepsilon) = 0, \quad |\varepsilon| < b_2 \tag{14}$$

Analytically the boundary wave of Rayleigh may be written in the form

$$\left.\begin{aligned} \varphi_1 &= \operatorname{Re}\left[G_1 f_1\left(pt \pm \tfrac{p}{\varepsilon} x + i\sqrt{\tfrac{p^2}{\varepsilon^2} - \tfrac{p^2}{a_1^2}}\, y\right)\right] \\ \psi_1 &= \operatorname{Re}\left[G_2 f_1\left(pt + \tfrac{p}{\varepsilon} x + i\sqrt{\tfrac{p^2}{\varepsilon^2} - \tfrac{p^2}{b_1^2}}\, y\right)\right] \end{aligned}\right\} y > 0$$

$$\left.\begin{aligned} \varphi_2 &= \operatorname{Re}\left[G_3 f_1\left(pt \pm \tfrac{p}{\varepsilon} x - i\sqrt{\tfrac{p^2}{\varepsilon^2} - \tfrac{p^2}{a_2^2}}\, y\right)\right] \\ \psi_2 &= \operatorname{Re}\left[G_4 f_1\left(pt \pm \tfrac{p}{\varepsilon} x - i\sqrt{\tfrac{p^2}{\varepsilon^2} - \tfrac{p^2}{b_2^2}}\, y\right)\right] \end{aligned}\right\} y < 0 \tag{15}$$

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ASTRONOMY

ON THE ORIGIN OF COMETS

By O. J. SCHMIDT, Member of the Academy

The theory of the planets' origin put forward by the author naturally leabs to the following ideas as to the origin of the comets. The latter, like the meteorites now observed, are r e m n a n t s o f a m e t e o r i c s w a r m that was once formed around the Sun, by capture, out a cloud of dark matter, which the Sun had crossed on its way in the central plane of the Galaxy. The main mass of this swarm has been used up in the formation of the planets, but a part has remained and continued revolving around the Sun.

This theory can be verified in the following way. The event that has given origin to the comets must also have influenced the spatial distribution of their orbits. This should be true at least in the case of long-period («parabolic») comets. Relative to the comets of short period such expectations seem unwarranted because these comets have performed too many revolutions and have travelled too long in the vicinity of planets, subject to perturbations which must have considerably altered their original orbits. Nor should we expect too much from the study of long-period comets. They, too, might have changed their orbits since the time when the solar system was formed. So it is only traces of the initial distribution, more or less obliterated in the course of time, that we can hope to reveal.

The nuclei of the comets in the form of accumulation of meteorites can have already existed in the galactic cloud or formed in the swarm by mutual capture. Some such accumulations are in progress of formation even now, while others are in a stage of disintegration. There is also room for impacts as an additional factor in a redistribution of meteoric matter, such as collisions between meteorites and asteroids, according to the theory advanced by Orlov (2).

1. In the present paper some laws are deduced from the theory of the origin of comets, which the distribution of the orbits of long-period comets should satisfy with the above-stated qualification. These laws will also be checked by facts. It appears that the laws of orbit distribution become rather conspicuous when reduced to galactic coordinates instead of to the customarily employed ecliptic coordinates. The recalculation to galactic coordinates of the orbit poles of all known long-period comets (with a semi-axis major larger than 100 astr. units) was done by A. I. Parenago under the guidance of P. P. Parenago to both of whom I feel much indebted. The well-known catalogue by Yamamoto (3) was used as a basis, and the additional information from the current literature on comets up to 1944 inclusive was supplied by an unpublished catalogue compiled by P. P. Parenago, of the Sternberg Astronomical Institute. The total number of comets with a period of more than 1000 years was found to be 383.

2. The system of coordinates we are to use has the Sun for its centre. The X-axis is defined as the direction which the Sun's velocity had at the moment when the meteorites were captured and the Sun was crossing the central plane of the Galaxy. The Y-axis is drawn perpendicular to this direction and parallel to the

plane of the Galaxy. It is supposed to run towards the centre of the Galaxy. The Z-axis is perpendicular to the XY-plane and directed northward. In an earlier work devoted to the distribution of the binary star orbits (4) I obtained the following formulae for this system of coordinates. If α, β, γ be the angles formed with these coordinate axes by the velocity which the meteorite had prior to capture, we will obtain angular coefficients for the direction to p e r i h e l i o n (1) and for the p o l e of the relative orbit formed (2)

$$\cos A_1 = 0, \quad \cos B_1 = -\frac{\cos\gamma}{\sin\alpha}, \quad \cos C_1 = \frac{\cos\beta}{\sin\alpha} \tag{1}$$

$$\cos A = \frac{\sin\alpha}{\lambda}, \quad \cos B = \frac{\cos\beta\sqrt{\lambda^2 - \sin^2\alpha}}{\lambda\sin\alpha}, \quad \cos\Gamma = \frac{\cos\gamma\sqrt{\lambda^2 - \sin^2\alpha}}{\lambda\sin\alpha} \tag{2}$$

where λ is the ratio of the relative velocity of the meteorite to the velocity of the Sun. I also introduced in the above-mentioned work the quantities ξ and η by means of the formulae $\alpha = \lambda\xi$, $90° - \beta = \lambda\eta$. The values of these quantities are considered to be equiprobable in the region of their definition, *viz.* $0 \leqslant \xi \leqslant 1$, $-\xi \leqslant \eta \leqslant \xi$. Proceeding to ξ, η and making use of the smallness of the angles α and $90° - \beta$, we get

$$\cos A_1 = 0, \quad \cos B_1 = -\frac{\sqrt{\xi^2 - \eta^2}}{\xi}, \quad \cos C_1 = \frac{\eta}{\xi} \tag{3}$$

$$\cos A = \xi, \quad \cos B = \frac{\eta\sqrt{1 - \xi^2}}{\xi}, \quad \cos\Gamma = \frac{\sqrt{\xi^2 - \eta^2}\sqrt{1 - \xi^2}}{\xi} \tag{4}$$

As the orbit of the Sun is but slightly inclined to the galactic plane, we may assume as an approximation that the Z-axis coincides with the direction to the p o l e o f t h e G a l a x y. Then the X-axis will lie somewhere in the galactic plane, and so will also the Y-axis. The directions of these axes can be determined by our theory.

3. For the angle Γ the distribution function has already been derived in the above-cited paper. Replacing Γ by $90° - b$, we can write the function then obtained in the form

$$f_1(b) = \sin 2b \ln \operatorname{ctg}\frac{b}{2} \tag{5}$$

This is the function of distribution of the galactic l a t i t u d e of orbit poles. Let now θ be the galactic longitude of the pole as measured from the X-axis. Employing the same method as in (4) for the calculation of the probability, we shall obtain the distribution function for the l o n g i t u d e

$$F(b) = \frac{1}{\cos^2\theta}\left[\frac{\pi}{2} - \theta\right] - \operatorname{tg}\theta \tag{6}$$

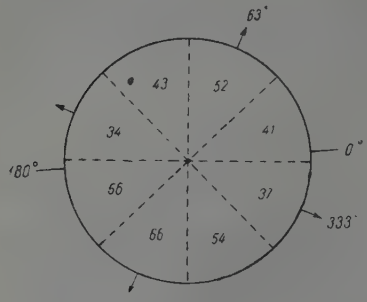
The function is determined for the first quadrant, to which the other can be reduced. This function vanishes in the limit for $0 \to \pi/2$ and is equal to $\pi/2$ for $\theta = 0$.

It will be obvious that the orbit poles exhibit an unmistakable tendency to group themselves near the XZ-plane and avoid the Y-axis. In other words, the orbital planes tend to cross each other about the Y-axis. If therefore, for some reason or other, meteorites accumulate into p l a n e t s, the plane of the planetary orbits will pass through the Y-axis. Now this axis has been stated to lie in the galactic plane, hence it is the l i n e o f n o d e s o f t h e e c l i p t i c i n t h e G a l a x y. Here the ecliptic is representative of all planetary orbit planes on account of their small inclination to each other.

This, indeed, is a most important result owing to which it is possible even now, after milliards of years have elapsed, to find the position of the Y-axis in the heavens. From the relations between the ecliptic and galactic coordinates we have for the line of nodes

$$\lambda = 86°.9 \quad \text{and} \quad \lambda = 266°.9, \quad \beta = 0$$
$$l = 153\ .2 \quad \text{and} \quad l = 333\ .2, \quad b = 0$$

That, in accordance with the theoretical forecast, the line of nodes of the ecliptic in the Galaxy is the line of minimum concentration of poles' projections, while the line perpendicular to it and containing the projection of the ecliptic pole onto the plane of the Galaxy is the line of maximum concentration, may be clearly seen from the adjoined figure. In it the galactic plane is divided into eight 45°-sectors, arranged symmetrically with respect to these lines and marked each with the number of comet orbit poles whose projections fall within the respective sector (out of a total number of 383 comets).



Also resorting to statistical data for comets we can obtain the direction of the least and greatest concentration by finding such galactic longitudes u that $\sum\cos^2(l_i-u)$ should become a maximum or a minimum. Here the sum extends over all long-period comets, and the l_i are the galactic longitudes of the poles. The solution is

$$\operatorname{tg} 2u = \frac{\sum \sin 2l_i}{\sum \cos 2l_i}$$

From factual data on the 383 comets we have

$$u = 65^\circ.1 \quad \text{and} \quad 245^\circ.1 - \text{maximum}$$
$$u = 155.1 \quad \text{and} \quad 335.1 - \text{minimum}$$

The direction of the minimum deviates from the above-mentioned node by less than 2°.

4. Let us compare the actual distribution of the longitudes and latitudes of the poles with the theoretical distribution as given by formulae (5) and (6).

Starting from the X-axis (taken approximately for $l=63°$) we divide the galactic plane into 30°-sectors, and superpose the quadrants upon each other without changing the position of the axes. The result of this procedure is given in Table 1.

Table 1

Longitude	Number	Actual %	Theoretical %
0—30°	158	41	60
31—60	127	33	30
61—90	98	26	10

The table reveals a qualitative analogy between the actual and theoretical distribution, though the course is less sharply expressed in the former than in the latter, which was to be expected, indeed.

In Table 2 are brought together the respective data on the galactic latitudes of the poles for all the 383 comets, as computed by Parenago. The data for the positive values of the latitude have been summed up with the data for the respective negative values.

For the latitudes the agreement between the theory and the data based on observation is even closer than for longitudes. Especially striking is the coincidence in the 20—29° interval where the theory predicts a maximum and the factual data really show it.

Table 2

Latitude	Number	Actual %	Theoretical %
0 — 9°	53	13.8	8.9
10 — 19	54	14.1	17.5
20 — 29	62	16.2	20.0
30 — 39	61	15.9	18.8
40 — 49	52	13.6	15.3
50 — 59	43	11.2	10.7
60 — 69	31	8.1	6.1
70 — 79	17	4.5	2.4
80 — 90	10	2.6	0.3

5. I know of only two works devoted to spatial distribution of orbits of the long-period comets: Svedstrup [5] in 1883 examined the directions to the perihelion. In 1924 Oppenheim [6] repeated his calculations on a more abundant material (306 long-period comets) and added to them computations of the distribution of the poles. Both employed the method of ellipsoid distribution, as used in stellar statistics. For the perihelia Oppenheim obtained the directions of the main axes of the ellipsoid as follows

$$\lambda_1 = 87.3 \qquad \beta_1 = -3.0$$
$$\lambda_2 = 182.4 \qquad \beta_2 = -59.8$$
$$\lambda_3 = 175.5 \qquad \beta_3 = +30.0$$

The third axis he found to be directed to the pole of the Galaxy (our Z-axis), and the first he approached to the vertex of the preferential motion of the stars. The fact that this axis (which is our Y-axis) very accurately points to the mutual node of the ecliptic and the Galaxy has escaped his attention.

On our theory it may easily be inferred from formulae (3) that X, Y, and Z must be the main axes of the ellipsoid. So here again the facts are in accord with the theory.

The study of the distribution of poles by means of the ellipsoid method will be carried out elsewhere in the light of our theory in connexion with the conditions determining the position of the ecliptic in space.

6. To sum up, it can be said that the here advanced theory of the origin of comets (at least of those with long periods) has led to the establishment of laws which have found support in factual data. Some of the laws derived from observation (the rôle of the line of nodes of the ecliptic and the Galaxy, distribution with reference to galactic longitudes and latitudes) were not noticed until they were predicted by the theory.

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BIOPHYSICS

DARK ADAPTATION OF THE EYE FOR INFRA-RED, VISIBLE AND ULTRA-VIOLET RADIATIONS

By N. I. PINEGIN

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It has been reported in literature that dark adaptation for rods proceeds at a much lower rate than for cones ([1-3]). From this essential difference it is possible to conclude which of the two apparatus, the photopic or the scotopic one, is responsible for the work of the eye under definite conditions.

The author was confronted by this question when measuring the absolute threshold sensitivity of the eye in the ultra-violet and infra-red part of the spectrum ([4,5]). Goodeve, Lythgoe and Schneider ([6]) investigated dark adaptation of the eye for $\lambda = 365$ mμ and for $\lambda = 546$ mμ. In both cases the authors have obtained analogous curves of the usual shape due to the superposition of scotopic and photopic adaptation curves. No measurements of adaptation for the infra-red are available.

In the present note is described an attempt to investigate the dark adaptation of the eye for infra-red, visible and ultra-violet radiations. Dark adaptation tests were preceded by a 5-minute exposure of the two eyes to white light. The brightnesses of the screen were $7.74 \cdot 10^{-1}$ sb and $3.4 \cdot 10^{-2}$ sb. The angle subtended at the eye was 60°. The sensitivity of the dark adapted eye was measured for 900, 800, 655, 579—577, 546, 365 and 313 mμ. Stray light was eliminated from the monochromatic radiations observed with the aid of filters assorted for this special purpose.

When determining the sensitivity of the eye to radiations of 900, 800 and 655 mμ a 15×0.2 mm monochromator slit was used as a stimulus, which was observed by means of an eye-piece with $\times 9$. The observer was invited to look straight, so that the image of the stimulus hit *fovea centralis*. When measuring the sensitivity of the eye to 579—577. 546. 365 and 313 mμ radiations, the angular size of a round stimulus was about 20 . The chromatic aberration of the eye was eliminated with the aid of negative quartz lenses of a suitable power.

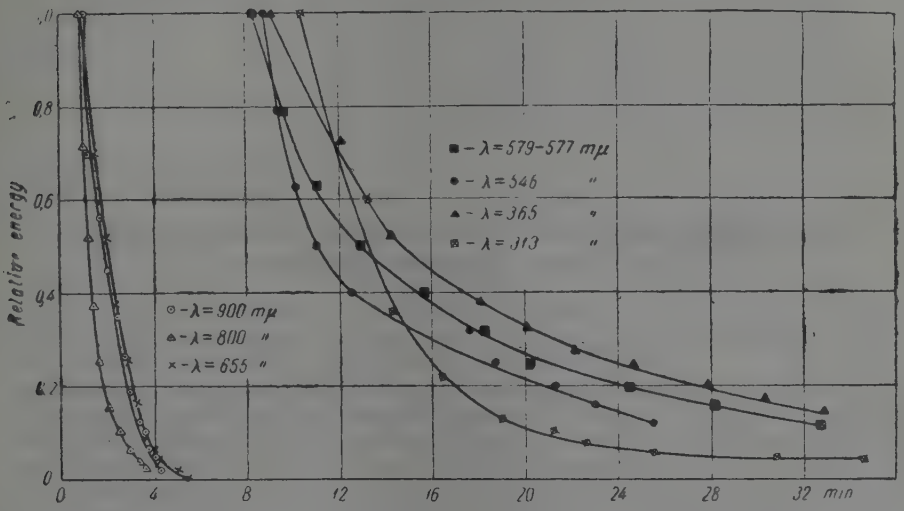
For each of the radiations in the region 579—313 mμ listed above the initial intensity was so chosen as to secure the work of achromatic, *i. e.* scotopic vision alone. Under these conditions the stimulus was observed in a reflex way by the most sensitive spot of the periphery of the retina.

In the course of dark adaptation the thresholds were determined with the aid of special photometric wedges. The results here reported have been obtained with three observers endowed with normal vision. As the run of the adaptation curves was found to be similar for the three observers, those obtained for one of them are reproduced here to serve as an example. Along the axis of the abscissae is plotted the time in minutes; along the axis of ordinates, the thresholds in relative units. The curves were obtained after the respective eye had been exposed to light of a $7.74 \cdot 10^{-1}$ sb intensity.

The figure displays a material difference between the length of the run of dark adaptation in the infra-red and red part of the spectrum, on the one hand, and the yellow-green and ultra-violet part of the spectrum, on the other hand. In the for-

mer case the curves attain the stationary threshold of eye sensitivity in the dark much sooner than they do in the latter case.

The group of curves obtained for $\lambda = 900$, 800 and 655 mμ illustrates a rapid run of the dark adaptation process. Unlike this, the group of curves plotted for $\lambda = 579$,



577, 546, 365 and 313 mμ depict a dark adaptation proceeding at a much lower rate. Thus, the beginning of the adaptation curves for $\lambda = 900$, 800 and 655 mμ corresponds in time to approximately 1 minute, and the end, to 4—5 minutes of the experiment. In contrast to this, the beginning of the adaptation curves for $\lambda = 579$, 577, 546, 365 and 313 mμ corresponds to 8—10 minutes, while their end corresponds to some 25—35 minutes of the experiment. The beginning of the adaptation curves in the $\lambda = 579$—313 mμ part of the spectrum is thus found to be shifted to no more than a half of the time observed for the end of the curves obtained in the $\lambda = 900$—655 mμ part of the spectrum.

Analogous relations have been observed for curves obtained after a preliminary exposure of the observers's eye to white light of an intensity of $3.4 \cdot 10^{-2}$ sb. The only difference was that with a lower intensity of preliminary lighting the curves of dark adaptation begin and end earlier than in the case of a greater brightness.

As a result one is brought to conclude that behind the material differences in the duration of the dark adaptation process studied by the author stands the work of two apparatus of the retina. In the red and infra-red spectrum, up to $\lambda = 655$ mμ at least, we have only photopic vision. In the ultra-violet, as well as in the visible spectrum, scotopic vision alone has been investigated, although photopic vision is also at work in this case.

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TECHNICAL PHYSICS

THE GENERATION OF HIGH-POWER ELECTRIC OSCILLATIONS BY A LOW PRESSURE DISCHARGE

By B. L. GRANOVSKY and T. A. SUETIN

(Communicated by N. D. Papalexi, Member of the Academy, 9. I. 1945)

1. An electric discharge in highly rarefied gases can be used to generate undamped electric oscilllations in various frequency ranges (see, for example [1,2]). Thus, undamped oscillations of varying amplitude, shape and frequency depending upon the conditions of the discharge, may arise in an ordinary discharge tube fed by a direct current even if the external circuit is strictly aperiodic. However, the power of these oscillations and the efficiency coefficient of the tube as a generator are low, owing to the fact that the current strength in the discharge fluctuates between two values other than zero without change of sign (oscillations of the first kind).

It is known that to get a good efficiency the current at the minimum must be reduced to zero (oscillations of the second kind), *i. e.* there must be complete interruption of the current. Such oscillations may originate in a tube having a narrow neck in the path of the discharge [3,4]. Kleinwächter [3] obtained undamped electric oscillations with a tube in which the anode region was separated from the cathode region by a metal diaphragm with a narrow opening. However, he was not able to obtain oscillations of any appreciable power, owing to the fact that an intermediate cathode spot frequently appeared on the diaphragm (cascade burning of the arc), which prevented interruption of the current.

We may get around this difficulty by preparing a diaphragm from a dielectric, for then it can no longer act as an intermediate electrode of the discharge. If this is done, undamped electric oscillations may be obtained whose power is limited only by the possibility of heat loss from the tube and the diaphragm.

Discharge tubes of different design with diaphragms made of dielectrics* were prepared, tubes having a liquid mercury cathode and an incandescent oxide cathode, a screening collar around the cathode or around the anode, and various cooling systems. All of them worked and could be used to generate oscillations of appreciable power.

The appearance of the oscillations is accompanied by a number of visual phenomena in the tube: light emission spreads throughout the tube and changes in colour; the anode heats up markedly under a comparatively small mean discharge current and with an abrupt rise in the mean voltage at the tube.

2. A quantitative study of the oscillations in a stenotrone was made with a cathode ray oscillograph, an electronic voltmeter and a wave-meter.

The frequency, amplitude and shape of the generated oscillations are affected by: a) the geometrical shape of the tube, b) the conditions of the discharge (nature and density of the gas, current value), and c) the make-up of the external circuit.

* To gas discharge tubes that generate electric oscillations owing to a tight neck in the discharge tube, we have given the name stenotrone (στενος, narrow).

The simplest case is when the external circuit contains only a source of electromotive force $\mathcal{E}$ and a resistance R. Fig. 1 gives a curve of the maximum and minimum values of the current strength in the tube measured with an

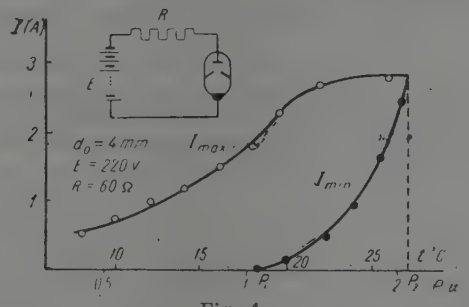

Fig. 1

oscillograph against a varying pressure of mercury vapour (the temperature of condensation of mercury is plotted along the axis of abscissae).

From the figure it is seen that under the given conditions (an aperture diameter, d_0, of 4 mm, $E=220$ V, $R=60$ Ω) right up to $t=18°$C ($p=p_1\cong 1\mu$ Hg) $I_{min}=0$, *i. e.* oscillations of the second kind occur, which are much different in shape from sinusoidal oscillations.

With a rise in temperature of the water bath (*i. e.* with an increase in p), I_{max} increases, and hence the amplitude of the oscillations also increases ($A=\frac{1}{2}I_{max}-I_{min}$). When $18°<t<26°$C, $I_{min}\neq 0$; interruption of the current

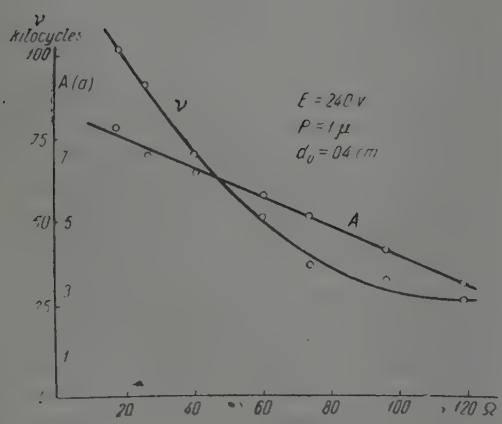

Fig. 2

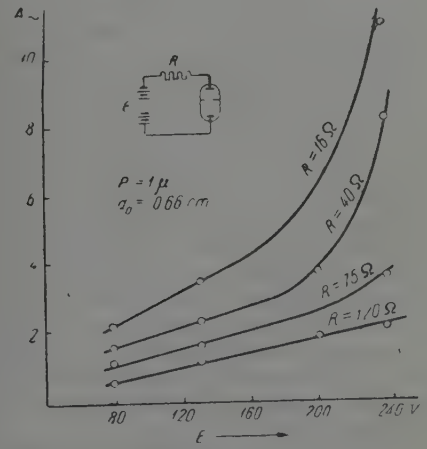

Fig. 3

is incomplete, *i. e.* oscillations of the first kind occur. The amplitude of the oscillations diminishes with an increase in p. Finally, when $t>26°$ C ($p\geq p_2\cong 2\mu$ Hg), $I_{min}=I_{max}$, the oscillations cease and the discharge become stationary.

The pressure p_1, at which the oscillations of the second kind pass over into oscillations of the first kind, depends upon the resistance of the circuit R: with a decrease in R p_1 increases. The same may be said about the pressure p_2, at which the oscillations cease completely.

In the case of oscillations of the first kind the smaller is R at a given p, the smaller I_{min} and the greater I_{max} and, hence, the greater the amplitude of the oscillations as well. At a certain R the current breaks completely (oscillations of the second kind set in). However, even with a further decrease in R, I_{max} still continues to mount, and hence the amplitude of the oscillations continues to increase as well (Fig. 2).

The frequency of the oscillations observed varied between 15 000 and 100 000 cycles. At these high frequencies, frequencies of the acoustic range are frequently superimposed—the coil in the anode circuit and the filaments of the indicator lamps «hum» distinctly. The fundamental frequency ν is a decreasing function of R (Fig. 2) and an increasing function of p.

An increase in the electromotive force of the battery intensifies the oscillations. In Fig. 3 we have plotted a family of curves expressing the dependence of the amplitude of the oscillations upon E at different values of R.

The area of the aperture in the diaphragm S essentially affects the generated oscillations: the larger the opening, the greater I_{max}. The increase in S of the aperture may be compensated by a corresponding decrease in the vapour pressure p.

The position of the diaphragm in the discharge also plays an important rôle. Experiments with tubes having a movable diaphragm have shown that as the latter approaches the anode, the value of the current at which the arc breaks increases; furthermore, there is such a position for the diaphragm at which we were not successful in obtaining complete interruption of the current (*i. e.* in obtaining oscillations of the second kind), notwithstanding the fact that the limiting resistance R was decreased to 4Ω. This phenomenon may be of significance for designing mercury rectifiers.

3. The observed phenomena fit in well with the conceptions developed by Hull([4]), Klarfeld ([6]), Tonks ([5]) and Kleinwächter ([3]) that the interruption of the current in an arc is due to rarefaction of the gas in the necks of the discharge.

The oscillations of the second kind may be easily conceived as alternate periodic breaks and ignitions of the discharge occurring in alternation, owing to the periodic decrease in concentration of the gas in the neck under the action of the passing current to a value insufficient to maintain the discharge and the reverse flow of gas into the neck, following interruption.

According to Hull and Elder, the current strength that causes interruption is proportional to the product of p and S. This law easily explains, for example, the fact that I_{max} is an increasing function of p and S, or the dependence of p_1 and R. It fails, however, to explain a number of observed relationships (for example, the dependence of I_{max} upon R at given p and S), and also the fact that the oscillations of two kinds may exist under the same working conditions of the discharge tube.

To explain these phenomena it is necessary to take into account that both the change in concentration of the gas in the neck and the change in current strength (development of the discharge) proceed at a finite velocity. The discharge current determines not the concentration of the molecules, but the rate of its change (dn/dt), and conversely: the degree of rarefaction of the gas in the neck affects the rate of the development of the discharge, *i. e.* the change in current value (dI/dt).

It is easy to see that I_{max} is determined by the operation of two competing processes: the process of increase in current in the circuit, and the process of rarefaction of the gas in the neck, and hence the value of I_{max} should depend upon R (the smaller R, the more quickly the current rate increases, and at the point of critical rarefaction the current may become very large).

These processes in the stenotrone are described by a system of differential equations of an order not lower than the second; that is why oscillations of the first kind become possible.

4. To utilize the oscillations generated by a stenotrone, the latter may be linked with the load by means of an inductive or capacitative coupling. In the case of an inductive coupling the primary winding of a high-frequency transformer is inserted into the anode circuit of the stenotrone, which it is well to tune to the frequency

of the generated oscillations by means of a variable capacity inserted in parallel. The load is inserted in the secondary winding of the transformer.

It is found possible to obtain from an operating stenotrone a highly useful oscillatory power P_∞ and a satisfactory efficiency η with a potential supply source of low voltage. Thus, a power P_∞ utilized as a load may be easily brought up to 1 kW and higher, while the efficiency computed as P_∞/P_0 increases up to 60—65 per cent for a source having an electromotive force of 240 V.

It is this latter circumstance that makes the stenotrone of technical interest as high-frequency generator, since, unlike the electronic valve, it requires no high voltage for feeding. It should also be noted that the stenotrone, as experiments have shown, may also operate in a circuit fed by alternating current.

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TECHNICAL PHYSICS

ON APPROXIMATE INTEGRATION OF VAN DER POHL'S EQUATION

By V. V. KAZAKEVICH

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In the present note a method of approximate investigation of non-linear systems is exposed, taking van der Pohl's equation as an example. This method enables us to determine the process of establishment, the form and the period of oscillations.

1. Consider the equation of van der Pohl

$$\ddot{x} - \mu(1-x^2)\dot{x} + x = 0 \tag{1}$$

Putting $x = y$, we write it in the form

$$\frac{dx}{dt} = y, \qquad \frac{dy}{dt} = \mu(1-x^2)\,y - x \tag{1a}$$

It may be easily shown that the motion around the equilibrium position $x = y = 0$ defined by the system (1a) may be described by two independent dynamic systems

$$\frac{dx}{dt} = y, \quad \frac{dy}{dt} = -\mu(1-x^2)\,|y|\,\operatorname{sgn} x - x; \tag{2a}$$

$$\frac{dx}{dt} = y, \quad \frac{dy}{dt} = \mu(1-x^2)\,|y|\,\operatorname{sgn} x - x \tag{2b}$$

the integral curves of which in the phase-plane are symmetric with respect to both axes of coordinates *

The system (2a) describes the motion in the second and fourth quadrant of the phase-plane (1a); the system (2b) in the first and third quadrant. In view of the symmetry of our integral curves we shall in the sequel call the systems of the form (2) quasi-conservative **.

Suppose that the equations of the integral curves (2a) and (2b) have been found. For the initial conditions $t = t_0$, $x_0 = a$, $y_0 = 0$ let them be

$$y_1 = y_1(x, a, 0), \quad y_2 = y_2(x, a, 0) \tag{3}$$

Put in (3) $x = 0$ and write

$$y_1(0, a, 0) = v_1(a), \quad y_2(0, a, 0) = v_2(a) \tag{3a}$$

* The case of a centre connected with symmetry with respect to one of the coordinate axes was considered by Poincaré [1] and Liapounov [2]. M. Almukhamedov [3] considered the case of a centre connected with symmetry with respect to an arbitrary straight line passing through the origin.

** In the most general case of a system of two differential equations of the first order the motion around every of its equilibrium positions may be described by four independent quasi-conservative systems.

The so obtained functions of the initial values (amplitudes) will be described in the sequel as determining functions. We shall consider the $v_i\,(i=1,2)$ in the fourth quadrants of the phase-planes.

If the initial amplitude a_1 is given, then subsequent amplitudes $a_2, \ldots, a_n, a_{n+1}, \ldots$ are determined by the equations

$$v_1(a_1)=v_2(a_2),\quad v_1(a_2)=v_2(a_3),\ \ldots,\ v_1(a_n)=v_2(a_{n+1}),\ \ldots$$

If there are stable periodical motions in the system (1a) (including equilibrium) and the initial value a_1 is chosen in the region of attraction of one of them, then the stationary amplitude is the limit of a sequence of amplitudes $a_1, a_2, \ldots, a_n, \ldots$ for $n\to\infty$.

The number of real roots of the equation

$$v_1(a)=v_2(a) \tag{4}$$

and their values determine the number and the values of the stationary amplitudes.

Let a^* be a real root of the equation (4); this stationary amplitude is stable, if we have the inequality

$$\left|\left(\frac{dv_1(a)}{da}\right)_{a=a^*}\right|<\left|\left(\frac{dv_2(a)}{da}\right)_{a=a^*}\right| \tag{5}$$

With the inverse sign of inequality the stationary amplitude will be unstable. The case of equality requires supplementary investigations.

2. To find approximate solutions of the systems (2) we substitute each of them with an auxiliary system constructed in a special way, and also quasi-conservative and oscillatory, which is obtained from the original system by means of a transformation varying continuously with x. Determining for each of the auxiliary systems the approximate expression for 1/4 of the oscillation period and returning to the original systems (2), we find their approximate solutions.

Let $x=a_0,\ y=0$ for $t=0$. Fixing $x=\bar{x},\ 0\leqslant\bar{x}\leqslant a_0$ and introducing new variables $\xi=x-\bar{x},\ \eta=y$, we consider instead of the system (2a) the system

$$\frac{d\xi}{dt}=\eta,\ \frac{d\eta}{dt}=\{-\mu[1-(|\xi|+\bar{x})^2]|\eta|-(|\xi|+\bar{x})\}\operatorname{sgn}\xi \tag{6}$$

It is easily seen that the time of movement from $x=a_0$ to $x=\bar{x}$ for the system (2a) coincides with that from $\xi=a_0-\bar{x}$ to $\xi=0$ for the system (6), i. e. is equal to $^1/_4$ of the oscillation period of (6) with the amplitude a_0-x.

We introduce now new variables by means of

$$\xi=(a-\bar{x})\cos(p_1t+\gamma),\qquad \eta=-p_1(a-\bar{x})\sin(p_1t+\gamma) \tag{7}$$

where a and γ are certain functions of time, and p_1 is a constant.

Instead of (6) we obtain the system

$$\frac{da}{dt}=\eta\cos u+\{\mu[1-(|\xi|+\bar{x})^2]|\eta|+(|\xi|+\bar{x})\}\frac{\sin u}{p_1}\operatorname{sgn}\xi. \tag{8}$$

$$\frac{du}{dt}=-\frac{\eta\sin u}{a-\bar{x}}+\{\mu[1-(|\xi|+\bar{x})^2]|\eta|+(|\xi|+\bar{x})\}\frac{\cos u}{p_1(a-\bar{x})}\operatorname{sgn}\xi$$

$$u=p_1t+\gamma$$

where ξ and η must be substituted from (7), and $u=p_1t+\gamma$.

Choose now p_1 in such a way that the mean value of du/dt for $^1/_4$ of an oscillation should be equal to p_1 (for $a-\bar{x}=\text{const},\ a=a_0$):

$$\frac{2}{\pi}\int_0^{\pi/2}\frac{du}{dt}\,du=p_1 \tag{9}$$

Computing this integral, we obtain a quadratic equation for p_1, the solution of which yields

$$p_1(a,\bar{x}) = \frac{\mu}{2\pi}\left[2-a^2-\frac{\bar{x}}{3}(2a-\bar{x})\right] + \\ + \sqrt{\frac{\mu^2}{4\pi^2}\left[2-a^2-\frac{\bar{x}}{3}(2a-\bar{x})\right]^2+\frac{4\bar{x}}{\pi(a-\bar{x})}+1}$$

Consequently, by (9) for the approximate value τ of $^1/_4$ of the period of oscillation for the system (6) we have $\tau = \pi/2p_1(a,\bar{x})$.

Making now $\bar{x}$ variable and replacing τ by t_1, $\bar{x}$ by x, we find the approximate solution of the system (2a) in the form

$$t_1 = \pi/2p_1(a,x) \tag{10a}$$

In a similar way we obtain the approximate solution for (2b)

$$t_2 = \frac{\pi}{2p_2(a,x)} \tag{10b}$$

where

$$p_2(a,x) = -\frac{\mu}{2\pi}\left[2-a^2-\frac{x}{3}(2a-x)\right] + \\ + \sqrt{\frac{\mu^2}{4\pi^2}\left[2-a^2-\frac{x}{3}(2a-x)\right]^2+\frac{4x}{\pi(a-x)}+1}$$

3. Putting in (10a) and (10b) $x=0$, we find the duration of every quarter of an oscillation

$$\tau_1 = \pi : 2\left\{\frac{\mu(2-a^2)}{2\pi}+\sqrt{\frac{\mu^2(2-a^2)^2}{4\pi^2}+1}\right\} \tag{11a}$$

$$\tau_2 = \pi : 2\left\{-\frac{\mu^2(2-a^2)}{2\pi}+\sqrt{\frac{\mu^2(2-a^2)^2}{4\pi^2}+1}\right\} \tag{11b}$$

The averaged half-period will be

$$T/2 = \tau_1+\tau_2 = \frac{1}{2}\sqrt{4\pi^2+\mu^2(2-a^2)^2} \tag{12}$$

The obtained formula establishes the approximate relation between the amplitude of an oscillation and its period. If the amplitude is approximately known, then (12) determines the period. For the equation of van der Pohl the value of the stationary amplitude is approximately equal to 2. Putting $a_{st} \approx 2$, we find

$$T_{st} = 2\sqrt{\pi^2+\mu^2} \tag{13}$$

If $\mu \ll 1$, then

$$T_{st} = 2\pi\left(1+\frac{\mu^2}{2\pi^2}+\cdots\right) \tag{13a}$$

If $\mu \gg 1$, then

$$T_{st} = 2\mu+\frac{\pi^2}{\mu^2}+\cdots \tag{13b}$$

In the following table are stated the values of τ_1, τ_2 and T_{st} computed by (13) for different μ.

μ	τ_1	τ_2	T_{st}	T (5)	T (4)
0	$\pi/2$	$\pi/2$	2π	2π	2π
1	2.15	1.15	6.6	6.6	7.2
5	5.45	0.45	11.8	11.3	—
10	10.2	0.25	20.9	18.6	20

It must be observed that with increasing μ the value of τ_1 increases rapidly, while that of τ_2 is rapidly diminished. It is of interest to compare (13) with the results already known.

For small μ Poincaré's method gives for the period of stationary oscillations the expression $T_{st} = 2\pi\left(1 + \frac{\mu^2}{16} + \cdots\right)$. For very large μ according to van der Pohl [4] $T_{st} = 1.618\ \mu$.

For arbitrary μ the following empirical formula, founded on graphical calculations, was proposed by Rensuke Usui [5] for want of a theoretical formula

$$T_{st} = 2\pi \sqrt[3]{\left[1 + \left(\frac{1.614\,\mu}{2\pi}\right)^{1.5}\right]^2}$$

The results of computation by this formula [5] are listed in the last but one column of the table alongside of the results found graphically by van der Pohl [4]. From the tabulated data our theoretical formula (13) appears to give sufficiently accurate results for arbitrary μ.

The expressions (11a), (11b), and (12) enable us to discover a number of subtle details in the mechanism of variation of the period and form of oscillations during the establishment process. For a increasing the half-period decreases from the value $T/2 = \sqrt{\pi^2 + \mu^2}$ for $a = 0$ to the value $T/2 = \pi$ for $a = \sqrt{2}$, and then monotonously increases with a. The form of an oscillation for $a < \sqrt{2}$ essentially differs from that for $a > \sqrt{2}$. For $a < \sqrt{2}$ the first quarter of the oscillation has a smaller duration than the second ($\tau_1 < \tau_2$); for $a > \sqrt{2}$ the case is reverse ($\tau_1 > \tau_2$). These last results being quite unexpected and at the same time important for the verification of the validity of the formulae (11a), (11b) and (12), it is desirable to compare these theoretical deductions with the results of direct (graphical or numerical) calculations.

Van der Pohl [4] gives the diagram of the process of establishment obtained by means of a twofold graphical integration of (1) for $\mu = 1$, which confirms (11a), (11b) and (12) quatitatively as well qualitatively *.

4. Differentiating with respect to time (10a) and (10b) and putting $x = 0$ we find the determining functions

$$v_i(a) = \frac{2}{\pi} \frac{1}{\left[\frac{d}{dx}\left(\frac{1}{p_i(a, x)}\right)\right]_{x=0}} \quad (i = 1, 2) \tag{14}$$

The determination of the establishment process may conveniently be carried out geometrically by constructing graphs of the determining functions in the plane $v - a$. The construction performed for $\mu = 1$ according to (14) is in fairly good coincidence with the graphical calculations of van der Pohl.

Investigation of the stability by means of (5) shows the unstability of the equilibrium and the stability of the limit cycle.

It may be remarked in conclusion that the method described is easy to modify for application to non-oscillatory systems.

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* Thus, the value of the ratio τ_1/τ_2 varies according to (11a) and (11b) from $\tau_1/\tau_2 = 0.52$ for $a = 0$ to $\tau_1/\tau_2 = 2.02$ for $a = 2.07$; according to van der Pohl's diagram τ_1/τ_2 varies from $\tau_1/\tau_2 = 0.45$ to $\tau_1/\tau_2 = 1.8$, respectively.

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TECHNICAL PHYSICS

THE ANALYSIS OF COMPLEX GAS MIXTURES WITH M. L. VEINGEROV'S OPTIC-ACOUSTICAL GAS ANALYSER

By S. M. LUCHIN

(Communicated by S. I. Vavilov, Member of the Academy, 5. I. 1945)

M. L. Veingerov (1) has suggested a new express method of gas analysis based on the ability of gases and vapours to emit sounds under the action of intermittent radiation.

This method was made the principle of an optic-acoustical gas analyser for the quantitative analysis of simple mixtures of known composition (2,3), which has been successfully applied in industry.

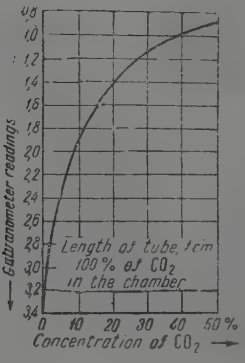

Fig. 1. CO_2 content in air

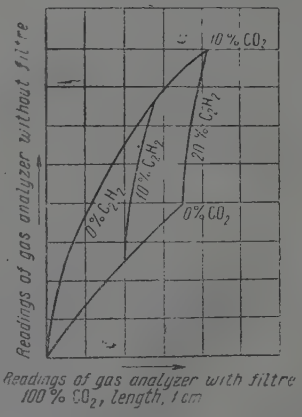

Fig. 2. C_2H_2 and CO_2 content in air

Veingerov recently designed a new model of the gas analyser in which the mixture for analysis does not flow through an optic-acoustical chamber, as in the previous apparatus, but through an absorption tube, placed between the chamber and the source of radiation. The chamber serves as a radiation receiver and is filled with the investigated component of the mixture, or with the mixture itself at some constant concentration. The mixture under investigation upon passing through the tube absorbs the radiation and in so doing decreases the amount of light entering the chamber, which leads to a decrease in the intensity of the sound.

With an increase of concentration in the mixture the intensity of the sound decreases.

The sensitivity of the apparatus can be appreciably increased by increasing the length of the absorption tube. The gas mixture may be passed through the tube at any velocity and still have a high temperature, which could not be done in the previous version of the apparatus. Moreover, in the latter it was necessary to isolate the optic-acoustical chamber during a determination to prevent outside noises from affecting the microphone. This does not have to be done in the new ap-

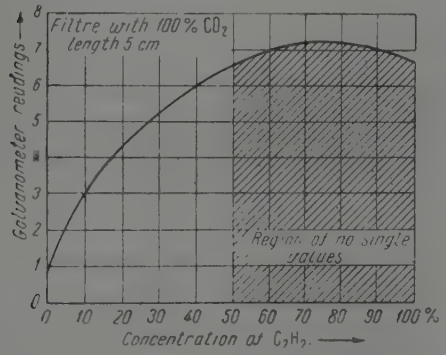

Fig. 3. Acetylene content in carbon dioxide

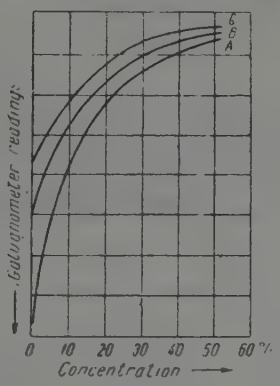

Fig. 4. *A*, *B*, *C* content in air

paratus, which allows continuous automatic control of the composition of gas mixture.

The calibration curve of the change in current strength at the output of the amplifier with the mixture concentration is shown in Fig. 1.

Attempts were made to apply the apparatus to the analysis of complex, multi-component gas or vapour mixtures; for this purpose Veingerov suggested the method of gas filters [1–3]. In 1940 this filter method was tested by Veingerov in collaboration with the author in the first model of the gas analyser and was found to increase quite adequately the precision of measurements at high concentration for single component mixtures [3]. In the case of complex mixtures, however, it did not always yield satisfactory results, owing to the complexity of the concentration diagram (Fig. 2), and sometimes owing to its ambiguity (Fig. 3).

Here we suggest a new method of complex mixture analysis effected with Veingerov's second model, which is free of these shortcomings.

Suppose that a complex gas or vapour mixture consisting of the components *A*, *B*, *C* is passed through the absorption tube. The task we set ourselves is to determine the concentration of each of those components. We fill the optic-acoustical chamber with one of the components gas, say *A*. In this case the chamber will respond only to the presence in the mixture of this component, since the gas (vapour) *A* filling it absorbs only the part of the radiation, corresponding to its absorption bands. The higher the concentration of *A* in the mixture, the less of the radiation, corresponding to its absorption bands enters the chamber, and the lower the current strength at the output of the amplifier. In this way a calibration curve for *A* is obtained, from which the concentration of this component may be found.

The chamber is then filled with another gas (vapour) *B*, its concentration is found from the corresponding calibration curve, and so on. The concentration diagram for such a mixture is shown in Fig. 4.

Practically it is more convenient to have as many chambers as components to be investigated.

In a determination the chamber with the desired component is placed in the path of the rays, and the pointer of the output instrument indicates its concentration.

In this method it is important that the absorption bands of the gases (vapours) under analysis should not overlap too much in the spectral region employed. When this condition is observed, both quantitative and qualitative analysis of multi-component mixtures becomes possible. Should the bands of some of the components strongly overlap, however, the latter can be analysed only with the help of monochromatic radiation ([4–6]).

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CHEMISTRY

BENZYLATION OF AMINO-DERIVATIVES OF NICOTINE

By J. L. GOLDFARB and M. S. KONDAKOVA

(Communicated by A. N. Nesmeyanov, Member of the Academy, 15. XI. 1944)

Our previous paper (1) dealt with results obtained by studying the action of alkyl iodides on α- and α′-amino-nicotines. The conclusion to which we were led by the discussion of the reported experiments can briefly be formulated as follows: the process of alkylation of these amines depends on the nature of the alkylating agent. This fact is brought into evidence by a change of the ratio between the amounts of Pl-iodoalkylate and the Py-substituted* compound when one alkylating agent is substituted with another. In this connexion it seemed interesting to investigate how distinctly would this phenomenon be expressed when the change in the nature of the alkylating or aralkylating agent consisted only in a substitution of one halogen with another. For this purpose we undertook a study of the action of benzyl chloride and benzyl iodide on the amines above-mentioned. The main results obtained in this study are exposed below.

With α′-amino-nicotine in ethyl alcohol or in acetone, benzyl iodide, as was to be expected considering the data on its reactive power, solely forms Pl-iodobenzylate (m. p. after crystallization from aqueous ethyl alcohol 190°). The action of benzyl iodide on the same amine, taken as a base, leads to the formation of a mixture of products, some of which are soluble in ether. Our attempts to isolate individual compounds from this mixture were, however, a failure. Nevertheless, the fact that products soluble in ether were formed suggests that the above-mentioned phenomenon will also take place when benzyl iodide is substituted with benzyl chloride which is a less powerful reagent. To get an idea as to the properties of the substitution products that would result from a reaction between benzyl chloride and α′-amino-nicotine, we started a new series of experiments in which this amine was taken as its monochlorohydrate. Accordingly, the influence of the pyrrolidine nitrogen atom upon the process was eliminated.

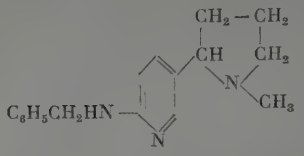
In these experiments the reaction product after treatment with an alkali proved all but completely soluble in benzene. By distillation *in vacuo* of this mixture a fraction was isolated which soon crystallized. After recrystallization from alcohol these crystals melted at 126° (picrate, m. p. 204—205°).

According to analysis data, this substance is a mono-benzyl derivative of α′-amino-nicotine. The same compound is obtained in good yields when α′-amino-nicotine is heated with a mixture of benzaldehyde and formic acid. Under the action of nitrous acid it forms a nitroso-derivative, m. p. 96°. Refluxing this derivative with diluted hydrochloric acid yields the initial base, m. p. 126°.

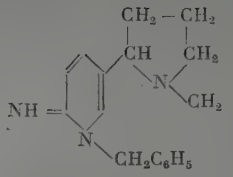
From this method of synthesis (2) and from the capacity of this compound to form a nitrosoamine, the structure of the substance melting at 126° appears to corres-

* Pl, pyrrolidine; Py, pyridine.

pond to formula (I) rather than (Ia).

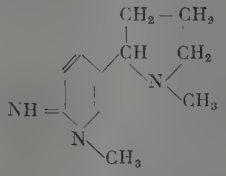


(I)



(Ia)

At first we were unable to decide whether the second structural isomer was or was not present in the mixture obtained from the action of benzylchloride on α′-amino-nicotine. Later, when we discovered that the product of methylation of α′-amino-nicotine of the structure (II) was precipitated by carbon dioxide, we turned again to this problem and applying the method of precipitation with carbon dioxide achieved the necessary results.



(II)

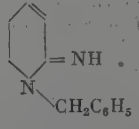
When CO_2 is passed through an etheral solution of the mixture formed in the benzylation of α′-amino-nicotine, a crystalline salt (probably the carbonate) is precipitated, while benzyl-α′-amino-nicotine, m. p. 125°, remains in the solution. From this salt it is possible by the usual method to isolate the base which, according to analysis, is the mono-benzyl derivative of α′-amino-nicotine. It has a colour while in the vapour phase and rapidly absorbs carbon dioxide when exposed to the air. All of this speaks in favour of the substance having the (Ia) structure. It seems probable that the property of forming salts under the action of CO_2 is peculiar to compounds having an alkylated pyridone-imino-group of atoms (III), whereas isomeric compounds having a group of atom (IV) do not enjoy this property.

(III)

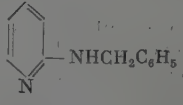
(IV)

(V)

In complete agreement with this conjecture is the fact that N-methyl-pyridone-imine (V) and benzyl-pyridone-imine (VI) are precipitated by carbon dioxide from their ether solutions, whereas benzyl-amino-pyridine (VII) is indifferent to the action of carbon dioxide.



(VI)

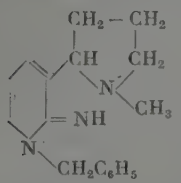


(VII)

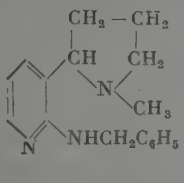
The reaction between benzyliodide and α-amino-nicotine proceeds less evenly than between the former and α′-amino-nicotine. In this case the addition of alkali to the mixture causes an oil to separate,from which but a very small amount of a substance can be extracted with ether. From the remaining oil it is only with difficulty that crystalline substance,the iodo-benzylate of α-amino-nicotine, m. p. 141°, is isolated. On the contrary, the reaction between α-amino-nicotine and benzyl chloride (in ethyl alcohol) leads to results that are easier to interpret. The product

formed in this case is almost completely extracted with benzene. When distilled *in vacuo*, part of α-amino-nicotine is recovered and the residue distils within a short temperature range. The purification of the main reaction product is achieved by converting it into the perchlorate and by crystallizing the latter from water.

The base isolated from this salt has a b. p. 216—128°/9 mm. It has colour in the vapour phase, and its composition is faithfully described by formula (VIII).



(VIII)



(VIIIa)

The analysis of its picrate (m. p. 204°) gives results that agree with computation.

In order to decide whether the product of benzylation of α-amino-nicotine contains the second isomer having the structure (VIIIa), the method of separating such isomers by means of carbon dioxide was used again. It was found however that carbon dioxide precipitated all the benzylation product as a crystalline carbonate.

The action of a mixture of benzaldehyde with formic acid upon α-amino-nicotine yields a product boiling *in vacuo* in a broad temperature range. Our attempts to isolate an individual compound were not immediately attended with success. Later, however, when this product was purified from formyl derivatives through boiling with diluted hydrochloric acid, we were able to isolate after repeated distillation a fraction boiling within a 3—4° range, still containing an admixture of the initial amine. The latter was removed by means of nitrous acid, and the benzylation product had b. p. 177—178°/4 mm.

Analysis data and determination of the molecular weight of this base, as well as the analysis of the picrate (m. p. 186°) speak in favour of its being the mono-benzyl derivative of α-amino-nicotine. It differs in its properties from the product obtained by the action of benzyl chloride on α-amino-nicotine. By analogy the structure (VIIIa) might be ascribed to this substance, but as the action of nitrous acid on this compound yields no nitroso-compound, we are not yet entitled to make this conclusion.

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CHEMISTRY

ADDITIVITY OF BOILING POINTS AND CHELATED BOND IN AROMATIC COMPOUNDS

By V. N. UFIMZEV

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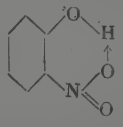
In the series of aromatic compounds in many cases considerable deviations are observed in the physical properties of ortho-substituted compounds. Thus, *o*-nitrophenol as compared to its isomers has a higher vapour pressure, boils at a lower temperature and may be distilled with steam. When molecular weight is determined by the cryoscopic method in non-polar solvents, a normal value is obtained for *o*-nitrophenol, and higher values, owing to molecular association, for its isomers. The solubility of *o*-nitrophenol in these solvents is higher, and its surface tension, specific weight and parachor are lower compared to its isomers.



In order to explain these phenomena, to which the name «ortho-effect» was given, Latimer and Rodebush (1) and later Sidgwick (2) introduced the concept of «hydrogen chelated bond», whereby the interaction of the ortho-substituting groups is supposed to lead to the formation of a coordination covalent bond in which hydrogen participates at the expense of the two unshared electrons of the oxygen atom. Other formulations of the mechanism of this phenomenon have been advanced by Sugden (3) and Pauling (4). Experimental studies (5) have demonstrated that hydrogen chelated bonds can be obtained by the interaction of atoms of the following types:

O – H ← O	O – H ← Hal
N – H ← O	N – H ← Hal
C – H ← O(?)	N – H ← N

Besides, it was pointed by Sidgwick (2) that the intra-molecular ring formed by the hydrogen chelated bond has to be either a 5- or 6-membered one; otherwise spatial strains interpose as a barrier to its existence.

According to the ideas expressed above for the formation of a hydrogen chelated bond the presence of two groups in the ortho-position is necessary: one containing hydrogen and the other being a donor of electrons for the formation of the bond. However, this is not always true, as for instance groups CH_3 (4,6), SH (7) as well as CHO and CH_2OH are incapable of the formation of hydrogen chelated bonds, notwithstanding the presence of a hydrogen atom.

The aldehyde group, containing a hydrogen atom directly bound with the carbon atom, forms no hydrogen chelated bond in spite of the presence of hydrogen, the boiling points of the isomeric benzaldehyde derivatives (Table 1) being at one and the same level. A substitution of the hydrogen atom by a methyl group leads to no essential change in these temperatures.

Table 1

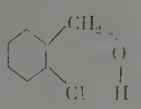
Boiling Points of Derivatives of Benzaldehyde and Acetophenone

Second substituent	First substituent: CHO-group, isomers			mm Hg	$COCH_3$-group, isomers			mm Hg
	o	*m*	*p*		*o*	*m*	*p*	
Cl	211	212	213		230	—	—	
Br	230	222	—		—	—	256	
NO_2	153	164	—	23	161	167	—	18
OCH_3	243	230	247		239	239	256	

Note. If not otherwise indicated, the pressure is 760 mm Hg.

The carbinol group containing OH forms no hydrogen chelated bond, because the existing small difference in the boiling points of the isomeric chloro-benzylic alcohols undergoes no change when hydrogen is substituted with the ethyl residue:

Substituents		Isomers: *o*	*m*	*p*
Cl	CH_2OH	227°	234°	234°
Cl	$CH_2OC_2H_5$	212°	219°	221°



In this case the requirement that the cycle be a 6-membered one is fulfilled and does not therefore hinder the formation of a hydrogen chelated bond. On the other hand, besides ortho-derivatives of phenol, ortho-derivatives of aniline and even of dimethyl aniline, containing no hydrogen directly bound with nitrogen (Table 2), have a much lower boiling point compared to their isomers, this being characteristic of compounds with a chelated bond.

Table 2

Boiling Points of Derivatives of Phenol, Anisol, Aniline and Dimethyl-aniline

Second substituent	First substituent: OH-group, isomers			OCH_3-group, isomers			NH_2-group, isomers			mm Hg	$N(CH_3)_2$-group, isomers			mm Hg
	o	*m*	*p*	*o*	*m*	*p*	*o*	*m*	*p*		*o*	*m*	*p*	
Cl	175	214	216	195	193	198	207	230	230		206	233	230	
Br	194	236	238	217	210	217	229	251	—		—	256	264	
											107	133	—	14
NO_2	215	—	295	272	258	260	211	230	260	95				
CHO	197	240	—	243	230	247					129	—	176	17
$COCH_3$	214	296	—	239	239	256	251	289	293		—	148	110	13
OH	245	281	286	205	244	243					199	266	—	
OCH_3	205	244	243	207	217	—	218	243	239		211	237	—	
NH_2				218	243	239	257	282	267		217	265	260	
$N(CH_3)_2$	199	266	—	211	237		217	265	260		216	266	260	

As a result of the cleavage of the chelated bond in derivatives of phenols which occurs when the hydrogen of the hydroxy-group is substituted by the methyl residue (with the exception of cases where the second substituent is hydroxy-, amino-, or dimethyl aminogroup) the compound boils at an increased temperature approaching the boiling point of its isomers.

All this shows that in the case of ortho-substituted derivatives of benzenes the hypothesis of the hydrogen chelated bond is insufficient to explain their properties or their physical constants and needs to be revised. This the author hopes to accomplish in the nearest future.

Conclusions. 1. Compounds having CH_3-, CHO-, CH_2OH- and SH-groups form no hydrogen chelated bond.

2. Derivatives of aniline and even of dimethyl-aniline (containing no hydrogen directly bound with nitrogen) behave similarly with the corresponding phenol derivatives so far as their boiling points are concerned.

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PHYSICAL CHEMISTRY

THE BURNING VELOCITY OF A BICKFORD FUSE AS RELATED TO PRESSURE

By K. K. ANDREEV

(Communicated by A. A. Skochinsky, Member of the Academy, 20. XI. 1944)

A Bickford or burning fuse consists of a black powder core tightly wrapped in layers of yarn which, in some makes, is impregnated with substances that render it waterproof. With a definite burning velocity of the order of 1 cm/sec. it provides the required delay in touching off the detonator, after its free end reaching out of the borehole has been ignited.

The burning velocity and its constancy are checked against accepted standards for the finished product by the fuse manufacturers. However, these tests are made at atmospheric pressure, whereas in practice the fuse is burned in a tamping, sometimes very thick, so that the combustion gases do not have free access to the atmosphere. As a result an appreciable pressure may develop in the narrow canal left by the burned out core due to the friction of the gases on its walls as they flow through, especially in the case of deep boreholes and also owing to clogging of the canal by solid combustion products.

But a rise in pressure is known to increase the burning velocity of explosives. Moreover, at an elevated pressure burning may turn into explosion. In either case the fuse may burn out much sooner than it is calculated to, setting the charge prematurely and injuring the operator.

Snelling and Cope [1] arrived at the conclusion that the increase in burning velocity with pressure is not constant between 1 and 6 kg/cm². At pressures from 2 kg/cm² and higher the burning velocity fluctuates within a much wider range than at atmospheric pressure, and from 4 kg/cm² upwards an explosive acceleration of the burning process is observed.

The object of the present work has been to study the character and velocity of the burning process at various pressures from the minimum, below which the process stops, to 30 kg/cm². Burning fuses manufactured in this country were experimented upon.

At pressures up to 2 kg/cm² the experiment was run in a glass-vessel connected to a container with constant pressure. At pressures above 2 kg/cm² a 1.3-litre steel bomb was used, connected with a 300-litre container and provided with a thick organic glass window, which made it possible to follow the burning process visually. In order that the end of the process might be distinguished with precision in the smoke-filled bomb, one end of the length of fuse ($l=25$ cm) was inserted into an opening in a rubber stopper at the bottom of which a bit of gunpowder was placed that produced a distinct flash.

At one atmosphere experiments were run in parallel in the bomb and outside it. The absence of a detectable difference in velocity indicates that the fuse does not heat up while burning in the bomb. Still less should we expect such a heating effect to take place at large pressures, *i. e.* at shorter burning time intervals.

The experiments at pressures up to 2.5 kg/cm² were carried out with two types of fuse: black and white; at the higher pressures only the second type was studied.

The results of the experiments with the white fuse are given in the accompanying table. They show that the burning velocity of a Bickford fuse increases with the pressure.

Burning Velocity of a White Fuse as a Function of the Pressure

Experiment No.	Pressure in kg/cm²	Velocity combustion v in cm/min.	lg p	lg v	Experiment No.	Pressure in kg/cm²	Velocity combustion v in cm/min.	lg p	lg v
1	0.279	26.5	−0.555	1.423	21	2.2	77.5	0.398	1.889
2	0.279	26.6	−0.555	1.425*	22	2.5	78.1	0.398	1.893
3	0.291	25.3	−0.536	1.403**	23	4.5	98.8	0.653	1.995
4	0.545	36.2	−0.263	1.559	24	4.45	97.7	0.648	1.991
5	0.545	35.1	−0.263	1.545	25	6.45	110.4	0.810	2.043
6	0.802	44.8	−0.096	1.651	26	6.45	107.5	0.810	2.031
7	0.802	45.8	−0.096	1.661	27	8.5	116.8	0.929	2.068
8	1.005	50.0	0.002	1.699	28	8.45	116.8	0.927	2.068
9	1.005	49.5	0.002	1.695	29	10.3	121.4	1.013	2.084
10	1.359	60.0	0.133	1.778	30	10.35	121.6	1.015	2.085
11	1.359	57.5	0.133	1.760	31	12.15	125.3	1.085	2.098
12	1.632	64.5	0.213	1.810	32	15.3	138.8	1.185	2.142
13	1.632	65.6	0.213	1.817	33	15.25	136.9	1.183	2.137
14	1.905	70.9	0.280	1.851	34	19.15	145.3	1.282	2.162
15	1.905	71.7	0.280	1.856	35	19.25	138.3	1.284	2.141
16	2.184	75.3	0.339	1.877	36	25.3	151.2	1.403	2.180
17	2.184	75.7	0.339	1.877	37	25.3	152.4	1.403	2.183
18	1.0	51.5	0.000	1.712***	38	31.3	164.1	1.496	2.215
19	1.0	55.3	0.000	1.743***	39	31.3	160.9	1.496	2.207
20	1.0	53.2	0.000	1.726****					

* Temperature 19—20° C.

** In tests 1—17 which were carried out in a glass tube the pressure fluctuated within ±0.005 kg/cm². In tests 18—39 which were carried out in a bomb the fluctuations ranged within ±0.05 kg/cm².

*** Burned in the open.

**** Burned in a bomb.

If for all homogeneous secondary explosives studied by the author (²) which in burning give no inorganic solid residue (trinitrotoluene, hexogen, tetryl, nitroglycol, pyroxilin and others) the velocity varies linearly with the pressure in accordance with the formula $v = A + Bp$, nevertheless for gunpowder the increase in velocity is by far less rapid. Analytically, the dependence of the burning velocity upon the pressure may be expressed by the relation $v = Bp^{\nu}$, where for a black burning fuse $B = 52.8$ and $\nu = 0.475$ when v is in cm/min. and p lies between 0.17 and 2.4 kg/cm²*. For white fuse $B = 50.9$ and $\nu = 0.528$ when p does not exceed 2 kg/cm², and $B = 72.6$ and $\nu = 0.238$ when p lies between 4 and 30 kg/cm².

From the established relations it follows that, compared to the other substances mentioned above, the burning velocity of gunpowder is least dependent upon the pressure**.

Thus, whereas the burning velocity of collodium cotton increases by 140 per cent upon raising the pressure from 10 to 30 kg/cm², the burning velocity of gunpowder increases by only 14 per cent in the same pressure range. Therefore the acce-

* At smaller pressures (0.016 and 0.102 kg/cm²) the fuse did not burn.

** We note that such a relation gives greater stability of burning in a semi-enclosed volume than the linear law. In this respect gunpowder and other substances subject to the same law in burning deserve unquestionable attention. In exactly the same way the slight dependence of the burning velocity upon the pressure characterizes gunpowder as the most suitable substance for various types of inhibitors operating under conditions of variable pressure, for which purpose it actually is widely used.

leration of the burning process in the borehole as a result of a possible increase in pressure will be relatively small, and from this point of view gunpowder is the most suitable substance for charging a slowly burning fuse. However, under such conditions as are met with in underwater blasting, the possibility of an acceleration in the burning process cannot be neglected, and it is necessary to increase the length of the fuse correspondingly.

Although the author's experiments were carried out in a considerably larger range of pressures than the above-mentioned experiments of Snelling and Cope, he did not detect the flash-like acceleration of the burning process, which the mentioned investigators observed beginning already with 4 atmospheres. It is possible that the acceleration they observed was due to heating up of the fuse ahead of the front, the influence of which, under certain conditions, I detected and eliminated from my experiments.

What has been said, however, concerns only the case of burning under a raised but constant pressure. In the case of burning under mounting pressure, as the investigations of the author on powdered secondary explosives have shown [3], the conditions for acceleration of the burning process are not restricted to the dependence of the burning velocity upon the pressure. For gunpowder, however, burning experiments carried out in narrow tubes (0.8 mm in diameter) open at one end have shown that a rise in pressure, which undoubtedly occurs as burning proceeds, does not cause the process to become explosive, but rather leads to its damping. Therefore for a Bickford fuse acceleration of burning to a point where explosion may occur seems highly unprobable. This is only possible in the case of a faulty fuse with a loosely packed core, which enables the burning process to propagate over the surface of the powder grains.

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BIOCHEMISTRY

ON THE EFFECT OF CERTAIN INHIBITORS OF METABOLISM UPON MOBILITY, RESPIRATION AND GLYCOLYSIS OF SPERMATOZOA

By I. I. IVANOV, K. I. KANYGHINA and M. V. UMANSKAYA

(Communicated by K. I. Skrjabin, Member of the Academy, 10. IX. 1944)

One of the authors of the present article has already shown (1) that the spermatozoa of mammals may use for their movements the energy released either in the process of respiration (oxidation of non-carbohydrate substances (2)), or in the process of anaerobic glycolysis.

If we are to assume that the reserves of phosphorous compounds (ATF and creatine phosphate) are actually playing in the mechanism of mobility of the sperm cell a part similar to that they are supposed to play in the mechanism of muscle contraction, the spermatozoon must be rather poor in these store substances. This is evidenced by the fact that in the case of a simultaneous blocking of both respiration and glycolysis the sperm cells are rapidly deprived of their mobility (2). These observations have been confirmed by Lardy and Phillips (3,4) without a reference to the above-cited work, it is true.

After the metabolism of spermatozoa had been studied in its general features, it appeared possible to pass to the investigation of the separate chemical conversions which supply the sperm cell with the energy necessary for its mobility. The influence of malonate which is a specific inhibitor of succindehydrase, upon the mobility and respiration of spermatozoa, has been studied by Ivanov (5), and Lardy and Phillips (4). Ivanov failed to reveal any inhibitory influence of malonate upon the mobility and respiration of bull spermatozoa obtained from the testicle annex; the American writers have found, however, that the respiration of ejaculated sperms is strongly inhibited by the effect of malonate.

According to Lardy and Phillips, the sperms acted upon by malonate (0.01 mol.) do not conserve their mobility unless kept in media containing glucose, *i. e.* under such conditions as enable them to use up the energy set free in the process of glycolysis.

In order to make out the cause underlying these contradictory phenomena which might well be due to the use of different test-objects (spermatozoa from the annex in the one case, and ejaculated spermatozoa, in the other) we undertook experiments involving both of them. We studied also the effect of malonate upon sperms intoxicated with monobromacetate, *i. e.* under conditions entirely precluding the possibility of the glycolysis energy being utilized by the spermatozoa for their mobility. Special care was invariably taken that the magnitudes of the pH and the osmotic pressure of the solution be always brought to their optimal values (pH = 7.4—7.6; Δt = 0.60—0.64°C). The results of several experiments out of the numerous tests carried out by us are summarized in Tables 1 and 2.

Table 1

Effect of Malonate upon Mobility of Ejaculated Bull Spermatozoa Separated from Excreta of Sex-Gland Annexes by Washing in NaCl-Phosphate Solution

Composition	Time after addition of malonate, in minutes, and mobility of spermatozoa									
	0	10	15	30	45	75	165	375	400	1000
NaCl-phosphate solution	good	good	good	good	good	good	good	good	satisf.	poor
NaCl-phosphate solution+0.2% of malonate . .	»	»	»	»	»	»	»	»	»	very poor
NaCl-phosphate solution+NaCN (0.001 mol.) . .	»	poor	slight* fluct.	—	—	—	—	—	—	—

* Upon addition of glucose the mobility was restored.

Table 2

Mobility of Ejaculated Bull Spermatozoa Washed in 0.9% NaCl Solution as Affected by Malonate in the Presence of Monobromacetate (pH=7.6; Δt=0.62°C); t = 25°C

Composition of medium	Time after addition of malonate, in minutes, and mobility of spermatozoa						
	0	10	15	35	95	125	420
Monobromacetate (0.0005 mol.)	good	good	good	good	good	satisf.	slight fluct.
Monobromacetate (0.0005 mol.)+malonate (0.2%) .	»	»	»	»	»	»	»
Monobromacetate (0.0005 mol.)+NaCN (0.001 mol.)	»	poor	immob.	immob.	immob.	immob.	—

As may be inferred from the data contained in Tables 1 and 2, the effect of malonate upon the mobility of spermatozoa differs sharply from that of the cyanide. In fact, in media devoid of sugar, or else in solutions poisoned with monobromacetate, the mobility of spermatozoa was observed to cease, as a rule, some 5—10 minutes after the cyanide had been added to it. Unlike this, in media containing malonate instead of cyanide, the mobility of sperm cells persisted for about the same period of time as in the control tests which did not contain this inhibitor of succindehydrase. Only in the case where use was made of solutions prepared by neutralizing malonic acid with sodium instead of NaOH, did we occasionally observe an effect similar to that described by Lardy and Phillips.

However, the decrease in the mobility of spermatozoa, which was found to set in some 45—60 minutes after the addition of malonate, was undoubtedly due to an excess of bicarbonate in the solution, as established by means of special control tests. The excess of bicarbonate is known sharply to reduce the life-time of mammalian spermatozoa, especially at body temperature (37° C). This effect of $NaHCO_3$ is to a considerable degree, or even entirely, eliminated with the aid of glucose.

According to our data, malonate, if prepared with due precautions, does practically fail to produce any effect upon the respiration of spermatozoa, or else, the respiration is reduced by no more than 10 per cent. The respiration of a pulp prepared with minced muscle tissue, however, was found to be reduced by 65—70 per cent.

Corresponding figures have already been quoted by one of the authors in his earlier paper ([5]).

We thus see that experiments on the study of the influence of malonate upon the metabolism and respiration of spermatozoa do not permit to claim for the oxidation system of succinic acid a decisive rôle in the respiration mechanism of sperm cells. On the contrary, these experiments make us answer the question rather in the negative. The subject is discussed in more detail in another paper by one of the present authors ([5]).

Ivanov and Burnasheva ([8]) have also considered the question as to the physiological significance of the glycolytic reduction-oxidation as the fundamental reaction of glycolysis, supplying the energy required by the sperm cell for fulfilling its mechanical function under anaerobic conditions. Whilst studying the behaviour of spermatozoa under anaerobic conditions in media containing glucose and the inhibitor of glycolysis, NaF, the above-cited authors were able to observe that the mobility of such spermatozoa poisoned with NaF is invariably restored under anaerobic conditions after pyruvate has been added to the medium. These data led them to the conclusion that the mobility of spermatozoa under anaerobic conditions in the presence of sugar is intimately connected with the reduction of pyruvic acid to lactic acid. It is this partial reaction of glycolysis that is evidently playing the most important physiological part.

Somewhat later it has been found by Lardy and Phillips ([7]) that although the formation by the spermatozoa of lactic acid does actually take place in the presence of glucose, NaF and pyruvic acid, the sperm cells do not preserve their mobility under these conditions.

At this juncture it should be pointed out that under anaerobic conditions, in solutions containing glucose and NaF, the spermatozoa are not restored to motion by an addition of pyruvate unless the concentration of NaF is a moderate one: it must be chosen in such a way as to prevent the cells from using up the glucose as energetic material, enabling them at the same time to maintain their mobility in the presence of air during at least 45—60 minutes. The initial concentration may be 0.01 mol., which is gradually to be reduced.

This circumstance seems to have been overlooked by Lardy and Phillips who worked with excessive concentrations of sodium fluoride, NaF.

Table 3

Influence of Pyruvic Acid upon Mobility of Spermatozoa and Formation by them of Lactic Acid under Anaerobic Conditions in the Presence of NaF. Spermatozoa Obtained from the Annex of Bull's Sex-Glands, t=25°C

Composition of medium	Time after the start of experiment, in minutes, and mobility of spermatozoa										Formation of lactic acid by the end of experiment in mm^3 of CO_2 (accord. to Warburg) *
	0	10	20	30	40	50	65	80	90	100	
Saline+glucose (0.2%)	good	good	good	good	good	good	good	good	good	good	144
Saline+glucose+ +NaF	»	poor	slight fluct.				abs.	abs.	abs.	abs.	4
Saline+glucose+ +NaF+pyruvate (0.2%)	»	good	good	good	good	good	poor	slight fluct. **			75

* All the figures are to be referred to 1 ml of the initial suspension of spermatozoa.
** A similar picture is observed under aerobic conditions.

Ivanov and Ivanova ([8]) have also investigated the effect of arsenate, which is a specific agent breaking the conjugated bond between the glycolytic redoxidation

and phosphorylation, upon the mobility of spermatozoa under anaerobic conditions. They succeeded in demonstrating that at a concentration of 0.01 mol., in the presence of glucose and under anaerobic conditions, arsenate does not affect appreciably the movement of sperm cells.

It has been found, however, that at a concentration of 0.01 mol. arsenate does rapidly (some 2—5 minutes after perfusion is started) arrest the work of an isolated frog heart poisoned with cyanide, *i. e.* of a heart which is operated by the energy of glycolysis.

It would, however, be premature to speak of an essential difference in the mechanism of contraction of the cardiac muscle and the tail of a spermatozoon merely on the ground of the observations reported above. No definite conclusions can be drawn until we evaluate the effect of arsenate upon the level of ATF in these objects and discard the possibility of an influence of arsenate upon the heart through a decrease in the concentration of calcium ions in the perfusate, as a result of binding a certain quantity of calcium by arsenate.

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GEOLOGY

ON THE GENESIS OF QUARTZ SANDS IN WESTERN SIBERIA

By V. P. KAZARINOV

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The Upper Cretaceous and the lower part of the Palaeogene in West Siberia are composed of a series of quartz sands, associated with white kaolin clays which do either enter the composition of the cement of the sands, or occur among them in separate bands. This peculiar series is widespread within the Chulym-Yenissei, near-Salair and Nenin-Chumysh depressions. By deep wells white quartz was penetrated in the West-Siberian lowland under a considerable thickness of Tertiary and Quaternary sediments. Quartz sands with kaolin are developed also in the European part of the Soviet Union, where they are traceable almost to the Balkan mountains.

It has been noted long since that quartz sands accumulated at definite periods of the life of the Earth. They are widely distributed in sediments of the Silurian, Carboniferous and Cretaceous. In other systems of the Palaeozoic and Mesocenozoic they do not play any significant part. On the basis of the theory of periodicity of sediment accumulation developed by him, Pustovalov (3) considers that at the beginning of great periods of sedimentation mainly polymikt, and at the end, mainly monomineral quartz sands are formed. At the beginning of a great period of sedimentation, corresponding to the initial phases of the epoch of a great revolution, large quantities of mineral detritus are supplied along the paths of transfer. Being transported under superficial conditions and subject to a constant action of weathering agents, this material was freed to an increasing degree from the alumino-silicate component, and in especially favourable cases was buried at the end of the period of sediment accumulation in the form of pure quartz sands.

Let us now consider the factual material available. We see that in West Siberia quartz sands take part in the composition of the products of Tertiary-Cretaceous weathering. The conditions of their formation and their place in the general section of the products of the weathering crust had been controlled by complex processes at the end of the Cimmerian cycle of tectogenesis. A stratigraphic tectonic scheme of the Tertiary-Cretaceous weathering crust of Western Siberia, based upon a large palaeofaunistic material, is presented in a simplified form in the following table.

The events at the close of the Cimmerian cycle of tectogenesis will be considered in succession. The Upper Cimmerian phase of tectogenesis brought about an uplifting of the Kuznetsk Alatau, Salair and some other areas. In the areas of depressions started the accumulation of Lower Cretaceous sediments, the upper horizons of which were composed of polymikt sands and argillites. At the close of the Lower Cretaceous time erosion brought the country to the level of peneplain, under conditions of which the weathering crust was formed. In the areas of applanation, subject to chemical weathering were friable sediments, mainly sands. An excellent diagnostic feature, showing that the kaolinization of sands occurred *in situ*, is

Stratigraphic and Tectonic Scheme of the Tertiary-Cretaceous Weathering Crust of West Siberia

Cycles of tectogenesis	Phases of tectogenesis	Phases of development of the weathering crust	Composition of the most characteristic products of the weathering crust		Age
			In areas of applanation* (Chulym-Yenissei, near-Salair, Nenin-Chumysh depressions)	In areas of deplanation* (Salair, Kolyvan-Tom folded zone, Kuznetzk Alatau)	
Alpine	Pyrenean		Redeposited products of the Tertiary-Cretaceous weathering crust		Oligocene
Cimmerian	Laramie / Austrian	Vaganov	Variegated clays of colloidal origin, bauxites, sedimentary ores, *etc.*		Eocene—Upper Cretaceous
			Eluvium of decomposition of quartz sands with a kaolin cement and of white clays	Eluvium of decomposition of sedimentary, metamorphic and magmatic rocks of the Palaeozoic and pre-Palaeozoic	
	Upper Cimmerian	Kiy	Variegated clays of colloidal origin, bauxites, sedimentary iron ores, *etc.*		Lower Cretaceous
			Eluvium of decomposition of friable rocks originally of a polymikt composition Slightly weathered and fresh polymikt sands and argillites	Eluvium of decomposition of sedimentary, metamorphic and magmatic rocks of the Palaeozoic and pre-Palaeozoic	
			Fresh rocks of a polymikt composition		Jurassic

* The name of applanation areas will be applied to areas brought to the level of peneplain by the accumulation of sediments. Deplanation areas are areas brought to the level of peneplain by means of a removal of material.

the preservation of the texture of the original rock—the alumino-silikate grains have passed into clay, retaining their original shape. Following Polynov (2), the author suggests for this peculiar eluvium of decomposition of friable sediments the name of neoeluvium. The analogon of neoeluvium in the areas of deplanation is the eluvium of decomposition of the country-rocks of the Palaeozoic and pre-Palaeozoic. A characteristic peculiarity of these clayey rocks is presented also by the relict textures of the original rocks.

At the expense of an intense removal of various elements (as colloids and as true solutions) from fresh rocks when they passed to a decomposition eluvium on the surface of the planed country, in its lake and bog water basins a horizon of peculiar products of colloidal origin was formed—fine, frequently marble-like mottled clays, bauxites, sedimentary iron ores, *etc.*

The Austrian phase of tectogenesis interrupted the formation of the weathering crust of the Kiy phase. New uplifts in the areas of deplanation took place. A renewed erosion destroyed the cover of the weathering crust of the areas raised. The products of erosion of the eluvium of decomposition in the form of kaolin clays and quartz sands were supplied to the applanation area, forming the Upper Cretaceous member of quartz sands and kaolins. The amplitude of uplifts of this time

was apparently small, not exceeding the thickness of the eluvium of decomposition (50—100 m). The uplifts occurred very slowly, the work of erosion operating in a sluggish fashion.

The sediments accumulated as a result of these processes are poorly fractioned, the grains are poorly rounded, the sands have generally a clayey cement.

The erosional activity revived by the Austrian phase of tectogenesis did not bring the country to the stage of peneplain, and the weathering crust did not complete this period of sediment accumulations. No Upper Cretaceous variegated clays of colloidal origin, bauxites, *etc.* are known in West Siberia. The erosional acitivity was revived again by a weakly pronounced Laramie phase of tectogenesis, which caused a continuation of the accumulation of quartz sands and kaolins and brought about an almost complete removal of the weathering crust in the areas of elevations.

The area of elevations of the Laramie phase, as compared to the area of elevations of the Austrian phase, was extended, owing to which the previously deposited Upper Cretaceous quartz sands with kaolin were subject to partial erosion. Therefore the upper horizons of the quartz sands, *e. g.* in the Chulym-Yenissei depression, in a general case are less agrillaceous, and the sand grains are rounded to a greater degree. Quite pure quartz sands occur there, too.

Towards the end of the Eocene the erosion brought the country to the level of peneplain, under the conditions of which the weathering crust of the Vaganov phase was formed. In the areas of applanation the series of quartz sands with kaolin were subject to chemical weathering. It should be noted that the quartz-kaolin series formed at the expense of erosion of the eluvium of decomposition of the deplanation areas contained a varying quantity of grains and pebbles of various rocks and minerals (feldspars, micas, *etc.*), fresh or weathered in a varying degree. The supply to the sediment of a fresh and incompletely kaolinized material was inevitable owing to the uneven, pocket-like shape of the bottom of the horizon of the eluvium of decompcsition, caused by the selective nature of the weathering process.

The supply to the sediment of fresh material was the greater, the deeper the erosion penetrated. This accounts for the presence in the upper horizons of the quartz sands of separate lenticular intercalations of sands of an almost polymikt composition, further on converted to typical neoeluvium.

Upon a kaolinization of the quartz-kaolin series, the grains and pebbles of the fresh rocks were converted into clays, retaining their original shape. The presence of pebbles and grains shows that the quartz-kaolin series represent a peculiar kind of neoeluvium.

In the deplanation zones the eluvium of decomposition was formed at the expense of fresh rocks of the Palaeozoic and pre-Palaeozoic. The eluvium of decomposition of the preceding phase of the weathering crust by the moment of formation of the Eocene weathering crust survived but in a few places on the Lower Cretaceous variegated beds in small depressions, in a belt of junction of the applanation and deplanation areas.

The variegated clays, as well as the bauxites, iron ores, *etc.* occasionally associated with them, and the underlying eluvium of decomposition are at present widespread within the areas where the weathering crust has been subject to insignificant erosion—the Kolyvan-Tom folded zone, South-West and North-West near-Salair zone, the Nenin-Chumysh, near-Salair and Chulyma-Yenissei depressions.

The Alpine cycle of tectogenesis interrupted in West Siberia the formation of the Tertiary-Cretaceous weathering crust. In the sediments of the post-Eocene time the products of the weathering crust occur in a redeposited condition, among sediments, bearing hardly any trace of chemical weathering. As a result of an erosion of the quartz-kaolin series in the Oligocene rather pure and well-rounded quartz sands were deposited. In the sediments beginning with the Miocene and higher up, no quartz-sands have been encountered in West Siberia.

The material presented above clearly shows that the quartz sands should be considered as a typical product of the weathering crust. As is known, associated with the weathering crust are numerous deposits of various mine-

rals — bauxites, sedimentary iron ores, manganese ores, fire-clays, *etc.* The sedimentary deposits of all these minerals are generally associated with quartz sands.

Such minerals as bauxite, pisolitic iron ores, *etc.*, being precipitated from water streams in the form of colloidal sediments, are localized among variegated clays composing the upper accumulative horizon of the weathering crust. It should be noted that the accumulation of colloidal sediments at the epoch of the weathering crust was always in a varying degree accompanied by mechanical sedimentation. Variegated clays often enclose therefore bands and lenses of quartz sands. It is also characteristic that the variegated clays and the sediments associated with them in the sandy fraction contain only minerals stable in the zone of chemical weathering, chiefly quartz. This is why the deposits of bauxites and sedimentary iron ores of the Lower Cretaceous and Eocene age, known in West Siberia, are associated with quartz sands, which, however, are in most cases highly argillaceous. The manganese ores of Chiaturi and Nikopol, the iron ores of Kerch, are known to be associated with quartz sands. The examples could easily be multiplied.

In analysing the materials on the Palaeozoic bauxites one is struck by their association with quartz sandstones. Thus, for instance, at the Pokrovsky deposit (western slope of the Urals) Meletschenko [1] recorded the occurrence of bauxites among ferruginized quartz sandstones. At the «Mezhevoy Log» deposit quartz sandstones lie at the base of the variegated beds. At the Maiski deposit (West Siberia, Salair) quartz sandstones are developed among limestones of the bottom and roof of the ore horizon. Several other examples could also be cited.

The association of the above-cited minerals with quartz sands is quite natural in the light of the accompanying genetic scheme. Quartz sands and their diagenized analogues—quartz sandstones—present reliable guiding features in the search for various minerals associated with the weathering crust.

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GEOLOGY

A CONTRIBUTION TO THE STRATIGRAPHY OF THE LOWER TRIASSIC OF THE SOUTH-USSURI COASTAL REGION (PRIMORIÉ)

By L. D. KIPARISSOVA

(Communicated by V. A. Obruchev, Member of the Academy, 1. VII. 1945)

The lower Triassic of the South-Ussuri Coastal region (Primorié), richly characterized by a diverse fauna, has received great attention on the part of those engaged in the study of the geology of this region. However, the results of their stratigraphic and palaeontological studies have not been summed up so as to form a single body of information. As a matter of fact, notwithstanding the abundance of fauna, no biostratigraphic section of the Ussuri Lower Triassic is as yet available; within it, three lithologically different horizons have been distinguished, and the greater part of the fauna was known to be confined to the middle horizon. Among those studying the Ussuri fauna, Diener (2) was the first to correlate the Lower Triassic deposits of the Coastal region on the basis of ammonites only to the lowermost part of the Himalayan Triassic with *Otoceras* beds; but later on Frech (3) raised their age considering them as equivalents of the Lower *Ceratites* limestone and *Ceratites* marls of the Salt Range (Solianoy Kriazh). Bittner (1) synchronized these beds, on the basis of lamellibranchs, with the variegated sandstone of Germany and with the Werfenian beds of the Alps.

On the basis of a study of lamellibranchs the present author also advanced the view that in the South-Ussuri Coastal region a complete section of the Lower Triassic must also be present (4). This suggestion has been confirmed by the study of cephalopods from the rich collection of V. D. Prinada of 1928 and from the collection of N. A. Nosdreyew of 1936. The volume of the brief characteristic of the stratigraphic section of the Lower Triassic of the South-Ussuri Coastal region given below permits to cite only a small part of the fauna, the genera and the species most characteristic of each horizon.

B a s a l b e d s. Resting upon the eroded surface of the Permian deposits and at places of Palaeozoic granites are conglomerates with pebbles mainly of granites and limestones with a Palaeozoic fauna. The sandstone cement of the conglomerate is occasionally found to contain a fauna of the Lower Triassic: *Ophiceras* (*Protophiceras*) *nicolai* Dien., *Grypoceras* cf. *hexagonale* Dien., *Gervillia exporrecta* Leps., *Myophoria laevigata* (Ziet.), *Anodontophora canalensis* (Cat.), *etc.* In the upper parts the conglomerates contain bands of sandstones similar to those composing the overlying beds of the section. The thickness of the basal beds varies from 10 to 75 m.

Meekoceras b e d s. These are represented by calcareous, sometimes quartz-feldspathic sandstones with bands of clay shales and with lenses of shell rocks. They include a rich and diverse fauna of cephalopods, pelecypods, brachiopods and, less frequently, of gastropods. According to observations by V. D. Prinada, confined to the lower part of the section of the *Meekoceras* beds are mainly lamellibranchs, among which prevail: *Velopecten minimus* Kipar., *V. bittneri* Kipar. and *Pseu-*

domonotis multiformas Bitt. In the middle part of the beds are contained chiefly ammonites belonging to the genera *Meekoceras*, *Proptychites*, *Ophiceras*, *Paranorites*, *Koninckites*, *Ussuria*, *Pseudosageceras*, *Gyronites*, *etc.* In the upper part lamellibranchs are again predominant, in particular *Pseudomonotis iwanowi* Bitt., and uniform brachiopods occur in large accumulations: *Terebratula margaritowi* Bitt. and *Spiriferina* sp.

The *Meekoceras* beds of the South-Ussuri Coastal region, from the *Meekoceras*, *Proptychites*, *Paranorites*, *Koninckites* and *Ophiceras* contained in them, may well be correlated to the Lower *Ceratites* limestone and to the *Ceratites* marls of the Salt Range (Solianoy Kriazh), or to the beds with *Meekoceras* of the Himalaya. The isolated specimens occurring in the *Meekoceras* beds and belonging to the genera *Xenoceltites*, *Kashmirites* and *Cordillerites*, known as yet in other countries from the higher zones of the Lower Triassic only, are represented by new species in the region considered.

In the case of a more careful method of collecting the fauna according to separate layers, the *Meekoceras* beds might possibly be subdivided. Their thickness attains 100 m, occasionally 200 m.

The name of «*Meekoceras*» beds is introduced by the author to replace that of «*Proptychites*» beds, sometimes encountered in the literature on the Triassic of the Ussuri region. Such a substitution seems to be rational, the genus *Meekoceras* being the guide genus for the determination of a part of the section of the Lower Triassic, and the beds with *Meekoceras* being traceable over many areas of development of the Lower Triassic deposits of the Earth.

Flemingites beds. Above the *Meekoceras* beds in the section appear thin-bedded sandstones with *Flemingites prynadai* Kipar. (in litt.) and *Proptychites latifimbriatus* de Kon. In addition to *Ceratites*, they contain large numbers of small pelecypods from the family *Nuculidae*: *Nucula goldfussi* Alb., *N. oviformis* Eck., *Ctenodonta elliptica* var. *praecursor* Frech, *etc.* The *Flemingites* beds seem to correspond in age to the *Ceratites* sandstone of the Salt Range, *Hedenstroemia* beds of the Himalaya and subzone with *Pseudosageceras multilobatum* of the *Meekoceras* zone of North America. The thickness of their beds and in general their more complete characteristics are unknown, since on the western coast of the Ussuri Bay the continuous section of Lower Triassic deposits is interrupted just at the sandstones with *Flemingites*, which have here an apparent thickness of 1 m only. In a more complete section of the Lower Triassic on the Russian Island no beds corresponding to these with regard to fauna have as yet been discovered.

Subcolumbites beds. On the Russian Island the *Meekoceras* beds are overlain by thin-bedded sandstones alternating with clay shales. In the lower part of their section some ammonites are encountered—*Tropiceltites inopinatus* Kipar. (in litt.), and 30 or 40 m higher in a limestone band up to 0.5 m thick a very rich and interesting fauna of *Ammonites* is contained: *Subcolumbites multiformis* Kipar. (in litt.), *Prosphingites globosus* Kipar. (in litt.), *Megaphillites immaturus* Kipar. (in litt.), *Paranannites gracilis* Kipar. (in litt.), *Pseudosageceras simplex* Kipar. (in litt.), *Grypoceras ussuriense* Kipar. (in litt.), *Orthoceras subcampanile* Kipar. (in litt.), *etc.*

Together with these there are abundant small *Palaeoneilo* (?) *prynadai* Kipar., and throughout the thickness of the *Subcolumbites* beds isolated specimens of the same family *Nuculidae* are encountered: *Nucula oviformis* Eck., *Ctenodonta elliptica* var. *praecursor* Frech, *Leda* sp. nov. Confined to the same beds are *Posidonia* aff. *tenuissima* Böhm., occurring in accumulations on bedding planes of some sandstones; and also vegetable remains—*Pleuromeia sternbergii* Münst. and *Equisetales* of the *Noealamites* type. The thickness of these beds is about 100 m.

The complex of genera of *Ammonites* of the *Subcolumbites* beds is almost entirely known from the *Subcolumbites* beds of Albania. The fauna of the two groups of beds seems to be younger than the fauna of the zone with *Columbites* of North America, since in the *Subcolumbites* beds of the region discussed here and in those of Albania are encountered genera widespread beginning with the Anisic stage of the Middle Triassic only (*Megaphyllites* and *Monophyllites*), and species closely approaching

Scheme of Correlation of Lower Triassic Deposits

Subdivision of the Lower Triassic after Spath (5)	South-Ussuri Coastal region	North-eastern Siberia	Salt Range	Himalaya	South-western China	Timor Islands	North America	Albania
Prohungari an	*Subcolumbites* beds	Olenek beds		Beds with *Prohungarites middlemissii*	Beds with *Beneckeia* and *Myophoria costata* Beds with *Pirolites*	Beds with *Palaeophyllites* and *Albanites*		Beds with *Subcolumbites*
Columbitan							Zone with *Columbites* Zone with *Tirolites*	
Owenitan			Upper *Ceratites* limestone	Zone with «*Sibirites*» *spiniger*	Beds with *Anasibirites* and *Kashmirites*	Beds with *Anasibirites* Beds with *Owenites*	*Meekoceras* zone: Subzone with *Anasibirites* Subzone with *Owenites* Subzone with *Pseudosageceras multilobatum*	
Flemingitan	*Flemingites* beds	Beds with *Hedenstroemia* and *Ophiceras*	*Ceratites* sandstone	Beds with *Hedenstroemia*	Beds with *Meekoceras*	Beds with *Meekoceras*		
	Meekoceras beds		*Ceratites* marls	Beds with *Meekoceras*				
Gyronitan			Lower *Ceratites* limestone	Beds with *Ophiceras*	Beds with *Ophiceras*			
Otoceratan	Basal beds		Shales with fossils	Beds with *Otoceras*				

the Anisic ones (*Paranannites suboviformis* Kipar., *Orthoceras subcampanile* Kipar.). Spath [5] also considers the fauna of the *Subcolumbites* beds of Albania as a later one than the fauna of the zone with *Columbites* of North America, but at the same time, in his opinion, it may not correspond to the fauna of the beds with *Prohungarites middlemissii* of Kashmir, nor to the equivalent fauna of the Olenek beds of the Arctics.

The present author believes that the *Subcolumbites* beds belong to the topmost portion of the Lower Triassic, since, as mentioned above, in them are encountered certain elements of the fauna characteristic of the Middle Triassic, and since in the South-Ussuri Coastal region they are directly overlain by Anisic deposits.

Like the Olenek beds, the *Subcolumbites* beds must correspond to the upper division of the stratigraphic scheme of Spath (1934), *viz.* to *Prohungaritan*.

A correlation of the section of the Lower Triassic of the South-Ussuri Coastal region with the corresponding deposits of other areas of their development and with the general biostratigraphic scheme of Spath is given in the accompanying table.

Within the South-Ussuri Coastal region the most complete section of Lower Triassic deposits is observed on the Russian Island and on the western coast of the Ussuri Bay. Basal and *Meekoceras* beds are, moreover, developed on the eastern coast of the Ussuri Bay, on the Putiatin Island, in the Abrek Bay of the Strelok Strait and on the western coast of the Amur Bay. Aside from the Russian Island, *Subcolumbites* beds may be present also on the eastern coast of the Ussuri Bay, on the basis of *Tropiceltites* cf. *inopinatus* Kipar. (in litt.) brought by A. I. Savchenko in 1936 from the Khanakhez River.

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GEOLOGY

ON THE RELATIONS BETWEEN THE BAIKAL AND THE SAYAN STRIKE IN THE ARCHAEAN OF EAST SIBERIA

CERTAIN CONCLUSIONS FROM AN ANALYSIS OF THE TECTONICS OF THE ALDAN PLATE

By E. M. LAZKO

(Communicated by V. A. Obruchev, Member of the Academy, 25. VI. 1945)

Chersky, a well-known explorer of East Siberia, established as early as the eighties of the last century that the Laurentian rocks of the oldest continent of inner Asia are characterized by two modes of strike of the folded structures: an east-north-eastern one, which he named the Baikal strike, and a west-north-western one, to which he gave the name of the Sayan strike (1). All the subsequent investigations inclusive those by Zverev on the Aldan plate (2) revealed the presence of these two predominant directions of parallel folds in the oldest rocks of East Siberia. The question as to the age relations between these two directions has still remained open.

From an analysis of a vast material the greatest authority in the geology of Siberia Obruchev describes the status of this problem in the following way: «Within a limited zone the displacement of one strike by another, more or less perpendicular, may be due to a plunge of the axes of folds, or brachyfolding, but for a vast area this explanation is inapplicable. One is to assume either a simultaneous folding in the shape of a series of arcs.. or one that took place at different time... In the latter case... it is more likely that the Baikal strike, the one most widespread on the «old crown» and most characteristic of it, is the older»(3).

The subject here considered is the problem of the age relations between the linear structures of the Archaean, developed over large areas; the variability of the strikes considered within a local area, as seen from the passage quoted above and from the data presented below, may have been caused by brachyfolding phenomena.

Extensive geological explorations of the Aldan plate, carried out during the few recent years, make it possible to approach the solution of the problem advanced at least for the eastern part of the Siberian platform.

The starting points in solving this problem are: in the first place, the determination of the age of the rocks of the Aldan complex; and secondly, new data on the tectonics of the Aldan plate.

As is known (4), the term of Aldan complex is applied to the complex of the most deep-seated Archaean granites, various crystalline schists, quartzites and marbles separated out by Korzhinsky as early as 1931 in connexion with the meridional crossing of the plate.

While investigating the valley of the Olekma River Arseniev succeeded in collecting highly valuable materials on the stratigraphy of this region; he also arrived at important conclusions concerning the structure of the whole East-Siberian crystalline massif. In the first place Arseniev established a division of the Ar-

chaean of the Olekma region into three members: as a result of a detailed correlation he arrived at the conclusion that the upper of the series separated out by him, the so-called Khana-Tingurcha series of quartzites, marbles, pyroxene and other crystalline schists, is of Upper Archaean age and is a stratigraphic analogue of the Aldan complex and of the rocks of the Upper Archaean of the Western near-Baikal area. An analysis of the distribution of the Archaean crystalline series of different age brought Arseniev to the conclusion that the «East-Siberian crystalline massif is the shape of a dome, the central part of which is located in the region of the Olekma-Vitim divide, where the lowermost portion of the Archaean has been exposed by erosions. East and westward this crystalline massif gradually plunges and is covered with a thick series of the Upper Archaean» ([5]). Arseniev's interpretation has recently been adopted by all geologists who have been studying the Aldan, in the first place by Pavlovsky.

On the other hand, the geological explorations of the last decade make it possible to draw an objective picture of the structure of the Aldan plate and to suggest a new interpretation of the observations of Zverev, according to which the strike prevailing in the south-western part of the plate is a Baikal strike.

From the geological data on the Central-Aldan region generalized by Bilibin and in part cited in his paper on post-Jurassic intrusions ([6]), and a review of all the materials on the central eastern parts of the Aldan plate recently made by Pavlovsky ([7]), it could safely be established that for the greater part of the plate the only persistent strike of the folded structures of the Archaean is a north-western, or Sayan one. Another important inference from these works is that «the predominant type of the folded zones of the Archaean consists of normal folds forming a system of linear folded structures» ([7]).

Summing up the materials of the geologists of Trust No. 13 and his personal observations, the author has arrived at the conclusion that in the western part of the plate the same linear folding is traceable, complicated in the extreme western parts of the plate by intensive phenomena of brachyfolding. Without dwelling at any length upon the causes of this phenomenon which will be the subject of a special study, let us consider it briefly.

In 1944 the geologists of Trust No. 13 discovered within the western part of the Aldan plate intensely dislocated rocks of Proterozoic (most probably, Lower Proterozoic) age, represented by quartzites and quartzite shales and clayey rocks and alevrites recrystallized to a varying degree.

The Proterozoic deposits have been crumpled into meridional folds, only the root portions of the synclinal folds squeezed amidst the Archaean basement being visible at present because of deep erosion.

The Proterozoic tectogenesis which produced the folded structure in the Proterozoic rocks, to a considerable extent affected also the older folded stuctures; as a result, the general plan of the Archaean structures in the western part of the plate is quite different from that in its central and eastern parts. The meridional strikes of the Archaean folds in the western part of the plate, entirely conformable to the strike of the Proterozoic folds, seem to have been caused by the rocks of the Archaean folded basement being also involved in the Proterozoic folding. The idea as to a superposition of folding processes of different age, as a result of which in the Olekma Archaean meridional folds of a complicated construction are observed, has also been advanced by Arseniev ([5]).

Between the areas of Archaean structures entirely re-oriented by Proterozoic tectogenesis in the western part of the Aldan plate, and those of undisturbed Archaean structures in its central part lies a transitionary area of development of brachyfolds which are frequently met with in the area between the Aldan and Chuga rivers. These folds also owe their origin to the action of Proterozoic tectogenesis, which resulted in a complication of the linear folds of this transitionary zone, formed during the Archaean, by sharp transversal bends. The transverse curvatures of the axes of the linear folds brought about the formation of complicating structures, in the form of brachyfolds and transverse fractures of the type of steep overthrusts. The alternation in the Aldan complex of roks of different competence

(quartzites and various crystalline schists) resulted in the fact that the quartzites which are particularly difficult to erode give well-shaped folds of a brachyal type, which are especially distinctly pronounced.

In the light of the above-said become clear also the age relations of the Baikal and Sayan strikes. The Sayan strikes which originated in the youngest Archaean rocks were the result of manifestations of the Upper Archaean tectogenic cycle, *i. e.* of a late fixed stage of Archaean folding. The Baikal strikes of linear folds, most characteristic of the «old crowns» where the oldest rocks of the Archaean are developed, are the result of a display of an earlier folding.

Thus, an analysis of the tectonic of the Aldan plate brings us to the conclusion that an interpretation which is in the agreement with the factual data is that given by Obruchev. According to this interpretation, the Baikal and the Sayan strikes originated at different periods, the former being earlier in age. In excellent agreement with the views of Obruchev is also the fact that the substitution of one strike by another perpendicular to it takes place in a local area of the western part of the plate and has been caused by brachyfolding.

The specific nature of this folding is well borne out by detailed mapping, and it is difficult to confuse its forms with common linear folds showing Baikal strike. The peculiarities of the structure of the extreme western part of the plate, as well as of the areas of the platform adjoining it from the west, composed of the lower members of the Archaean, are also evidence of the above statement.

Future detailed investigations in the various regions of the Siberian platform will show the possibility of extending the conclusions made to vaster territories and consequently of separating out two Archaean tectogenic cycles.

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PALAEONTOLOGY

DISCOVERY OF REPRESENTATIVES OF *TRIONYCHOIDEA* (*TESTUDINES, REPTILIA*) IN THE PLIOCENE OF THE UKRAINE

By L. I. KHOSATSKY

(Communicated by I. I. Schmalhausen, Member of the Academy, 10. VI. 1945)

Among the vast palaeontological material collected in the Ukraine as early as in 1908, the author discovered two fragments of bones which appeared to be fragments of plastra of peculiar fresh-water tortoises, the trionixes. The finding of remains of trionixes in the Ukraine is of particular interest, for the information available on representatives of the order reported for the USSR and the adjoining countries of Eastern Europe is rather scanty, indeed. The remains here described, together with some other data available, contribute to our knowledge of the palaeographic distribution of *Trionychoidea*, which differed greatly from the recent ones.

Trionyx sp.

Material. No. Kr. 1908; No. Nk. 1908. 253. Edge fragments of hyoplastra (hypoplastra?). The fragments are strongly rounded and mineralised. Preserved at the Zoological Institute of the Ukrainian Academy of Sciences.

Horizon. Upper Pliocene. Kuchurgan deposits.

Place of finding. Valley of the Kuchurgan River, Odessa province, the Ukrainian SSR.

In spite of the fragmentary character and the high degree of rounding of the remains, there can be no doubt as to their belonging to one of the middle parts of the plastron and, most probably, just to the hypoplastron. The cellular pattern of the external surface of the bones of the carapace, which is typical of *Trionychoidea*, is but poorly preserved in the remains here described; its traces are clearly detectable, however. The maximum dimensions of the bones are as follows: length, width and thickness of fragment No. Kr. 1908 are 76, 46 and 12 cm, respectively; for fragment No. Nk. 1908. 253 the respective figures are 35, 32 and 10 cm. By the dimensions and, in the first place, by the thickness of the bone fragments it should be concluded that they belonged to fully adult animals whose size was by no means inferior to that of the recent big *Trionyxes* from tropical regions.

Because of the fragmentary conditions of the material the determination of the genus cannot be considered as a definitive one, the more so as representatives of a closely related genus, *viz. Amyda* Oken., were also widespread in Europe during the Pliocene. The remains are therefore tentatively referred by the author to the genus *Trionyx* Geoffroy.

The place of occurrence of *Trionyx* sp. closest to the one referred to above has been reported by Sintzov (5); this are the bluish-dark clays in the exposures of Bolshoi Fontan, in the vicinity of Odessa. Simionescu (4) has also described *Trionyx*

sp. along with *Clemmys* sp. and *Testude* sp. from the Upper Pliocene of Bessarabia, village Bereshti. Prendel (2) has described a femur of *Trionyx* sp. from the Sarmatian formations in the environs of Sevastopol; according to informations supplied by Riabinin (3), A. A. Borissiak, Member of the Academy, had likewise found remains of *Trionyx* in these Sevastopol deposits. It should lastly be pointed out that among the collections of the Zoological Institute of the Academy of Sciences of the USSR is preserved a fragment of hypoplastron of *Trionyx*, labelled «*Trionyx* Kishinew. Nordmann». There is no reference as to any definite horizon from which this fragment had been obtained. It may well be assumed, however, that it was found in the Upper Neogene, the fauna of which was studied by Nordmann.

In Eichwald's «Palaeontology of Russia» (7) we find a remark to the effect that «... bones of tortoises had been discovered by Mr. Nordmann in the Molasse limestone of Bessarabia». It seems rather probable that these remains belonged to a *Trionyx*. As to the other cases of discoveries of tertiary *Trionyxes* within the area of the USSR, according to literary data (8), their remains are known from the Palaeogene of the Turgay province (Oligocene or Aquitanian strata of the Turgay, Min-Say), from the area near the Aral Sea (Palaeogene? of the Kizyl-Kumy) and the region of lake Zayssan (Palaeocene from the mounts Manrak, Kirghizian steppes).

From the Upper Cretaceous *Trionyxes* have been described for the region of Tashkent (Sary-Agach) and the Amur province. In none of the cases could the remains be referred to a definite genus and were merely determined as *Trionyx* sp.

Representatives of the order *Trionychoidea* had probably made their appearance as early as in the Middle Cretaceous, since their reliable remains were described from the Upper Cretaceous of America and Asia. In Europe they occur in the Palaeocene, disappearing in the Pliocene. At the present time *Trionyxes* occur in the Eastern zoogeographical province, in the south-east of Palaearctic, in equatorial Africa and in the plains of the USA, in the USSR the basin of the Amur River is the only place where they are encountered.

A historical survey of the palaeogeographical distribution of *Trionyxes* in Europe and America shows that, after having attained a rather considerable development during the Miocene, they underwent a strong retrogression by the close of Pliocene, which accounts for the sparse findings of their remains in the Pliocene. The extinction of *Trionyxes* during the Pliocene (this group had already once suffered retrogression during the Oligocene) may be put down to unfavourable climatic and other conditions that had prevailed at that period. This assumption is in good accord with geologic and palaeontological data according to which a continental steppe climate must have prevailed over a considerable part of Europe in the middle and towards the end of Pliocene. In the South of the Ukraine the landscape and climatic conditions underwent a change towards a higher aridity of the climate during the Pliocene; the area of the water reservoirs, of estuaries in particular (5), had been reduced, and there developed open areas of the type of the recent bush steppes of Africa (1).

Within the region where the finding of the remains of *Trionyx* here described was made, the landscape had probably been characterized towards the end of Neogene by the presence among shrubby steppes of rocky spots, with an uneven relief and stony talus. The aridity of the European climate during the Pliocene could not fail to affect the distribution of *Trionyxes* in an unfavourable way. This circumstance ultimately led to a total extinction of this group on the Europaen continent and in particular in the Ukraine, where they had been of no infrequent occurrence during earlier geological periods.

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CYTOLOGY

INVESTIGATION OF THE SMOOTH MUSCLE CELL AND ITS FIBRILS WITH THE AID OF MICROSURGICAL TECHNIQUE

By G. ROSKIN and N. VOLZHINA

(Communicated by L. A. Orbeli, Member of the Academy, 23. VII. 1945)

Very little is known regarding the physical properties of the chief components of the smooth muscle cell: its fibrils, the interfibrillar plasm and the cell membrane (the myolemma). Indeed, quite often even the very existence of fibrils as of definite individualized formations in the smooth muscle is questioned. Their inadequate visibility *in vivo* in many cases, the great difficulty of making them apparent in fixed and stained preparations, the lack of sufficiently reliable methods for their elective staining,—all these show how obscure the question has still remained.

It seems therefore quite natural to suppose that the smooth fibrils *in vivo* are but rows of definitely oriented particles which produce the impression of homogeneous filaments in fixed preparation, as the result of the coalescence of micellae due to the action of the mordant agents. It was surmised that by means of microsurgical experiments it might be possible to solve this question with some degree of certainty.

The observations were carried out with the aid of a Peterfi micromanipulator. In a number of cases the microsurgical experiment was combined with «dark field» (cardioid-condensor) microscopic examination, which permitted to record the changes in the aggregate condition of the cell components at different stages of the experiment. In a number of cases the operations on the muscle cell were supplemented by the action of certain reagents (20 per cent glucose solution, 0.05 per cent solutions of acetic acid or of potassium hydroxide, distilled water) on the cells which had been operated upon.

Cells of the obturator muscle of the mollusc *Anadonta* were used as the object of observation. These muscle cells, which had been used by many workers in their researches on the physiology and histology of smooth muscles, represent a most convenient object for microsurgical experimentation both owing to their considerable size and because their cells can easily be separated and isolated individually. Besides, in these cells the fibrils are readily discernible. The main difficulty encountered in working with these cells is due to the great toughness and elasticity of the connective-tissue intramuscular fasciae, making their tearing by the needles of the micromanipulator a rather strenuous task. After their isolation, the muscle fibres were stored in isotonic saline used for cold-blooded animals.

1. Experiments with Extension or Pricking of Muscle Fibres. Attempts to extend the muscle fibres permit to establish that they are but little extensible, unelastic and easily broken. Quite a different picture is observed in extending the muscle fibre (in this case the opposite ends of the long axis of the muscle fibre are preliminarily fixed). Under these conditions the muscle fibre can easily be distended transversely with the aid of microneedles, its breadth increasing 3—4 times.

It is essential to note that at the side where this distension is performed, the fibrils of the muscle fibre separate one from another, the distance between them increases and they show a characteristic arrangement (Fig. 1, *b*). If the distension of the muscle fibre is stopped, its original condition is restored.

If the microneedle is introduced into the plasm of the muscle cell, the regular character of the arrangement of its fibrils is disturbed (Fig. 1, *a*). In later expe-

Fig. 1

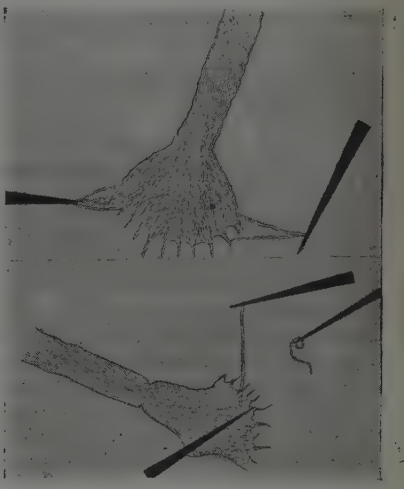

Fig. 2

riments we succeeded in extracting individual fine fibrils out of the muscle cell; these fibrils were, as it seemed, covered by a thin layer of interfibrillary plasm. Their apparence was that of fine homogeneous threads. They are distendable, elastic and curl up when broken (Fig. 2).

Experiments arranged for the study of the myolemma showed that the membrane of the muscle cell cannot be punctured unless some stress is applied; thus, the myolemma is certainly a solid stable formation. After the introduction of the needle into the muscle cell, changes may be observed in its cytoplasm: it loses its original transparency and becomes granular and more turbid than the neighbouring unimpaired portions of plasm. So far as this can be surmised on the ground of microsurgical experiments, the interfibrillary plasm of the muscle cell is viscous and ropy.

2. Phenomena Observed in Tearing Muscle Cells. By studying the phenomena observed at the site of rupture after tearing the muscle cell a more profound analysis of the properties of its individual components is made possible. The experiments with the tearing of muscle cells by ex- and distending it with needles were combined with a simultaneous action on them of the reagents enumerated above.

In observing the muscle cells in isotonic saline it may be seen that every fibril is surrounded by a thin layer of interfibrillary plasm. In the region of the rupture of the muscle cell the fibrils protrude from under the surrounding plasm in the shape of fine threads, solitary or distributed in more or less dense clusters. In transmitted light it is difficult to observe the individual fibrils separately, they can practically not be discerned from the surrounding plasm.

The picture which we observe if we use a paraboloid condensor is quite different. In this case the difference in the aggregate conditions of the fibrils and of the interfibrillary plasm is clearly apparent: in the dark field the interfibrillary plasm

is invisible, while the fibrils are clearly visible in the shape of light-blue homogeneous threads (Fig. 3).

These threads, protruding from the cell at the site of rupture, are swaying under the action of the currents of the surrounding med um. Muscle cells immersed in 0.05 per cent solutions of acetic acid or potassium hydroxide (in normal saline) did not show any changes in the structure of their fibrils in the course of observation for the period of one hour. The interfibrillary plasm in this case was changed, it became turbid, «dark», lost its transparency. The properties of the myolemma also changed: the cell membrane became less stable and could be punctured and torn by microneedles with greater ease.

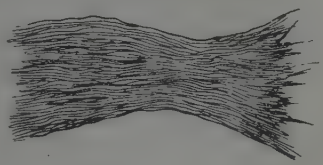

Fig. 3

Accordingly, it was possible to tear the envelope of muscle cells immersed in a 0.05 per cent solution of acetic acid with a relative ease, and seize individual fibrils or groups of fibrils united by interfibrillary plasm with the needles at the site of the break. In these experiments it proved possible to separate some of the fibrils from the neighbouring ones, while others could not be separated as if they were glued to each other.

Our next observations were made on muscle cells which had been kept for one hour in distilled water. In this medium the muscle cells swelled, their transversal diameter becoming nearly double its original size. The fibrils showed no visible changes in this case. Only the distance between them increased, apparently owing to the swelling of the interfibrillarly plasm.

The properties of the myolemma suffer a change under the action of distilled water: it becomes less solid and is impaired by a microneedle with greater ease.

Our last observations were carried on with muscle cells immersed in a 20 per cent solution of glucose, dissolved in normal saline. Under these conditions the transversal diameter of the muscle cell decreased a little. The fibrils approached one another closer. The interfibrillary plasm lost its transparency, it became «dark». It is probable that the transversal shrinking of the muscle cells in hypertonic glucose solutions is the result of changes in the interfibrillary plasm. No noticeable changes are observed in the fibrils of such muscle cells, except for their somewhat closer arrangement.

The above observations permit the following conclusions:

1. The smooth muscle cells of *Anadonta* have a characteristic membrane—a myolemma, endowed with a considerable solidity and a certain elasticity; under the influence of the reagents applied the myolemma becomes less resistant to tearing.

2. The interfibrillary plasm is optically homogeneous both in transmitted light and in the «dark field». It is a viscous, ropy plasm adhering to the fibrils and separated from them with difficulty; various interferences cause readily changes in the aggregate condition, which result in its becoming turbid if the preparations are examined in the «dark field». Swelling and unswelling of the interfibrillary plasm can be reproduced with relative ease.

3. The fibrils are homogeneous, definitely individualized formations; they can be removed from one another, dissected and isolated microsurgically. They prove to be resistant to the action of the reagents applied and, in distinction to the interfibrillary plasm, do not show any noticeable changes under the influence of these. The above results appear of special importance if they are considered in connexion with the earlier cytological observations of Roskin on the smooth muscle cells of various invertebrate and vertebrate animals.

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PLANT PHYSIOLOGY

ON THE EFFECT OF BIOGENIC STIMULANTS UPON THE DEVELOPMENT OF COTTON

By V. I. KOKUEV

(Communicated by A. A. Richter, Member of the Academy, 14. VII. 1945)

It has been shown by Filatov that when animal tissues are preserved at low temperature, *i. e.* under unfavourable conditions, the tissues that survive the treatment develop substances which, when introduced into the organism in the form of preserved tissue or extract, go to enhance its vitality. These substances whose nature is still obscure were termed by Filatov «conservation factors» or «biogenic stimulants» (1).

Filatov has proved, moreover, that plant tissues, when subjected to the action of low temperature and darkness, do likewise produce active substances which stimulate life activities in a way analogous to that of preserved animal tissue.

Experiments carried out by Filatov and others have shown that an extract prepared with leaves of aloe and applied to vegetable organisms (oats, Indian corn, *etc.*) has similarly a stimulating effect upon the growth and development of these crops (2).

At the Central Plant Breeding Station of the Institute for Cotton Industry an experiment was arranged in 1944 to study the effect of the extract of aloe upon the development and growth of cotton. The work was done under the guidance of V. P. Filatov and V. V. Skorodinskaya.

After being kept in a refrigerator at a temperature of + 2 to + 4 °C in full darkness during 10 days, the leaves of aloe were extracted with distilled water. Experiments involving the treatment of cotton seed with the extract of aloe were conducted after the following scheme:

Concentrations of the extract 1 : 10, 1 : 20, 1 : 40
Duration of the treatment 2, 4 and 6 days
Varieties used No. C 460 medium early, No. C 3173 early

The seeds were first soaked in running water and then subjected to the action of the extract for 2—6 days. The jars containing the seeds immersed in the extract were kept in the dark at a temperature of 2—6° during the time prescribed by the scheme of experiment. The amount of the extract was 10 cm^3 per 100 g of seed. As controls served seeds which had been kept for 24 hours in running water and seeds which were subjected to low temperature without treatment with extracts.

The experiment was started under field conditions on the 1st, 3rd and 5th of June. This is a very late date for sowing cotton at Tashkent.

Variety No. C 460 is the principal commercial variety of the Uzbek Soviet Socialist Republic; it belongs to the medium early group and is characterized by high productivity, high commercial value and by good technical qualities of the fibre. Variety No. C 3173 which belongs to the early maturing group is characterized by high productivity and resistance to wilt.

The aim of our experiments was to study the effect of the aloe extract upon the development of different cotton varieties with different duration of the treatment and different concentration of the extract. The seeds were treated with weak extracts only. The stimulation experiments were arranged in fourfold replication, each plot containing 20 plants. Records were made of developmental phases according to flower formation, as well as of the yield of raw fibre.

The observations have shown that under the influence of the «biogenic stimulants» obtained from the aloe extract, the plants developed at a higher rate, *i. e.* the phase of flowering was accelerated. Unfortunately, the terms of the dehiscence of bolls could not be duly recorded on account of the frosts that set in.

It should be noted that at the initial phase of development under field conditions all plots planted with stimulated seeds showed a certain depression of plants as compared with the controls. But later on the plants which had received the stimulating treatment long before the beginning of blooming started developing with a greater energy than the control plants. Nearly in all experimental series blooming was recorded earlier than in the controls (Table 1).

Table 1

Variety	Experimental series	Duration of treatment: 2 days — Blooming in days	2 days — Difference as compared with control	4 days — Blooming in days	4 days — Difference as compared with control	6 days — Blooming in days	6 days — Difference as compared with control
C 460	Control	63	—	64	—	68	—
	Cold	72	−1	63	−1	67	−1
	Aloe 1 : 10	63	0	64	0	66	−2
	» 1 : 20	62	−1	63	−1	65	−3
	» 1 : 40	61	−2	61	−3	64	−4
C 3173	Control	55	—	58	—	61	—
	Cold	55	0	56	−2	60	−1
	Aloe 1 : 10	54	−1	55	−3	60	−1
	» 1 : 20	56	+1	55	−3	58	−3
	» 1 : 40	55	0	55	−3	59	−2

From the above data it follows that in the majority of plants which had received the stimulating treatment blooming set in 1—4 days earlier than in the controls. The optimal concentrations proved to be in both varieties 1 : 20 and 1 : 40.

In the majority of series a better yield of raw cotton was also obtained (Table 2).

Table 2

Variety	Experimental series	Duration of treatment: 2 days — Yield of raw fibre, g per 1 plant	2 days — Difference as compared with control, %	4 days — Yield of raw fibre, g per 1 plant	4 days — Difference as compared with control, %	6 days — Yield of raw fibre, g per 1 plant	6 days — Difference as compared with control, %
C 460	Control	69	—	42	—	41	—
	Cold *	73	+ 5.8	41	− 2.4	36	−12.2
	Aloe 1 : 10	83	+20.3	52	+23.8	47	+14.7
	» 1 : 20	61	−11.6	54	+28.6	46	+12.2
	» 1 : 40	66	− 4.3	51	+21.4	50	+22.0
C 3173	Control	85	—	70	—	57	
	Cold *	86	+ 1.2	78	+11.4	49	−14
	Aloe 1 : 10	78	− 8.3	77	+10.0	71	+24.6
	» 1 : 20	112	+31.8	84	+20.0	84	+47.4
	» 1 : 40	92	+ 8.2	78	+11.4	78	+35.8

* By cold is understood the treatment with temperatures from +2° to +6° C.

From the above data one may see that with variety C 460 as a result of a 4-day-long treatment there was an increase in yield from 21 to 29 per cent in all experimental series.

With variety C 3173 an increased yield was obtained in all series (except one) within the limits 7—47 per cent. The best results, irrespective of the duration of treatment, were obtained with a 1 : 20 extract of aloe.

Two-day-long treatment of seeds gave indistinct results. Parallel tests of a sample treated only with cold has not shown any positive effect.

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PLANT PHYSIOLOGY

CONTENT OF INORGANIC P_2O_5 IN THE LEAVES AS AN INDEX OF A SUFFICIENT SUPPLY OF PHOSPHORUS AND OTHER ELEMENTS

By M. N. YAZVITSKY

(Communicated by A. A. Richter, Member of the Academy, 10. VII. 1945)

With the same amount of nutrient materials contained in the soil, the rate of their uptake by the plant may vary according not only to the concentration of the elements in questions, but also to a number of other agents (temperature of the soil, respiration of the roots, *etc.*). When using fertilizers, it is necessary therefore to take into consideration not only the amount of nutrient materials introduced into the soil in a form accessible to the plant, but also the intensity of their uptake by the plant. In order to control the plant's nutrition such an analytical method is necessary as would permit to find out whether it is possible at a given moment to stimulate the plant's growth by adding to the nutrient medium definite elements: it should also help to establish the mode of introducing fertilizers, which would result in an optimal assimilation of the nutrient substances by the plant.

Causing disturbances in the normal course of the synthesis the deficiency of some definite element may lead to the accumulation of intermediate products of synthesis or of other nutrient elements in the inorganic form. Certain writers have emphasized this circumstance [1–3].

It may seem that the introduction of the missing element into the medium would bring about a reduction of the amount of the intermediate products of synthesis (inclusive of nutrient elements in a modified form) at the expense of their transformation into more complex organic compounds. This was laid at the basis of our work undertaken with the aim of finding a method for controlling the nutrition of plants. With this aim in view we made use of the data on the variation in the inorganic P_2O_5 content under the influence of different conditions of nutrition.

As well known, the nutrient materials are absorbed by roots in an inorganic form. In particular, phosphorus enters the plant in the form of PO_4. A large part of the phosphoric acid reaches the fundamental organs of synthesis (leaves) in the form of ions [4]. Here it is combined with other nutrient elements and enters into complex organic compounds. This transformation requires a certain time. Hence at any given moment there should be a certain amount of phosphoric acid in the form of an inorganic compound. This amount will depend on the rate of synthesis, on the rate at which P_2O_5 enters the plant, and on the supply of other elements which are necessary for synthesis. By the intensity of the variation in the amount of inorganic phosphoric acid in the leaves caused by the introduction of other nutrient elements one may judge, therefore, whether the latter were in minimum, or were not.

To check this hypothesis we have, in the first place, to prove the existence of a direct correlation between the amount of phosphoric acid accessible to the plant in the soil and the amount of its inorganic form in the leaves, and, secondly, to find

out, whether the amount of inorganic P_2O_5 present in the leaves is reduced when the deficient nutrient elements are introduced into the medium.

For the analysis only such parts of the plant are to be taken where the synthetic process goes on at a very high rate. In the case of fruit-trees and small-fruit plants leaves should be selected from the growing shoots which have just attained their full size. In currant which was the main object of the present investigation, this will be the 4th or the 5th leaf from the growing cone.

The leaves are cut into small pieces, and an average sample weighing about 0.5g is taken. The latter is carefully ground with sand, 100 ml of 0.1 per cent solution of ammonium molybdate (in 0.4 *N* sulphuric acid) is added to the resulting pulp. The extract is stirred during 10 minutes and filtered. In the filtrate P_2O_5 is determined after Dénigé without combustion of the organic matter. Determination is made with the aid of a scale under the application of a comparator.

Although the method permits of determining along with inorganic phosphoric acid also a small fraction of the organic constituent, one may nevertheless provisorily regard the values obtained as relating to inorganic phosphoric acid.

In order to find out the relation between the amount of phosphorus present in the soil in a form accessible to the plant and the content of inorganic P_2O_5 in the leaves, a number of analyses were made of fruit-trees and small-fruit plants grown under different conditions of phosphoric nutrition and, as far as possible, under equal growing conditions. In all instances there was observed a clear direct connexion between the amount of inorganic phoshoric acid in the leaves and the amount of soluble phosphoric acid in the soil. The results obtained coincide with those by Filles and Mason ([5]) (Table 1).

Table 1

Doses of superphosphate in mg of P_2O_5 per 1 kg of soil	0	45	90	180	360
Inorganic P_2O_5, in leaves, %	100	178	195	217	329

In the field experiment the intensity of the uptake of P_2O_5 was recorded with the aid of the same method for different modes of introducing superphosphate.

Phosphoric fertilizers were introduced on the 3rd of June under 5-year-old currant shrubs, according to the three following methods: 1) on the surface of the soil with subsequent covering; 2) into four holes; 3) into 8 holes to a depth of 20 cm. After introducing the fertilizer the leaves were investigated for the content of inorganic P_2O_5 until the period of growth was over. One might expect *a priori* that in the case of superficial introduction of the superphosphate, phosphorus would be weaker absorbed by the plant and, consequently, less inorganic P_2O_5 would be accumulated in the leaves as compared to the case where superphosphate was introduced into holes. The analyses went to confirm this expectation.

Table 2

Content of Inorganic P_2O_5 in Leaves with Different Methods of Superphosphate Introduction (P_2O_5 in mg per 100 g of raw leaves)

Method of introducing superphosphate	Date of analyses				P_2O_5 on 16.VII, %
	11.VI	23.VI	5.VII	16.VII	
Superficially	18	10	20	22	100
Into four holes	22	26	32	46	209
Into eight holes	26	34	40	52	236

The second question to be investigated was the variation in the amount of inorganic P_2O_5 in the leaves with different doses of nitrogen and potassium fertilizers, the phosphoric nutrition remaining the same. Here, too, the above suggestions

were found to hold true. The introduction of deficient nitrogen or potassium resulted in a striking reduction in the amount of inorganic P_2O_5 in the leaves.

Table 3

Doses of N in mg per 1 kg of soil	0	45	90	180	360
Inorganic P_2O_5 in mg per 100 g of foliage	166	70	61	57	49
Same, %	100	42	37	34	30

In this experiment the introduction of nitrogen fertilizer went to increase the growth of currant by 50 per cent. A very clear effect was obtained in experiments on supplementary doses of nitrogen and potassium fertilizers.

In the first experimental series involving additional doses of nitrogen fertilizers, mineral fertilizers were introduced in spring; in the first series P and K only were introduced in spring, while nitrogen was given on June 4th. The soil for the experiment was meagre. Insufficient nitrogen nutrition caused in the plants of the second series a considerable accumulation of inorganic P_2O_5 towards the time when supplementary nitrogen fertilizer was introduced. The introduction of nitrogen brought about very rapidly a sharp reduction in the content of inorganic P_2O_5 in the leaves and had a favourable effect upon the growth of currant. Toward the time when supplementary fertilization was administered, the accretion of the plants of this experimental series was 54 per cent of that in the first series.

After the introduction of the additional nitrogen fertilizers, the plants of the second series have soon recovered, and towards autumn they could compare with those of the first series.

Table 4

Series	P_2O_5 in the leaves in mg				
	4.VI	10.VI	21.VI	29.VI	15.VII
NPK introduced in spring	52	40	36	28	36
PK in spring + N in summer	300	192	26	20	30

The introduction into the soil of the deficient potassium had, with regard to the content of inorganic P_2O_5 in the leaves, an effect similar to that of the nitrogen fertilizers.

Table 5

Series	P_2O_5 in mg				Total increment in cm	Total weight of foliage in g
	4.VI	10.VI	21.VI	29.VI		
NPK introduced in spring	52	40	36	28	160	276
NP in spring + K on 4.VI	110	56	48	32	184	291

Table 6.

Leaves taken from plants given the following relation of N : P : K	0 : 1 : 1	0 : 2 : 1	0 : 4 : 1	2 : 0 : 1	4 : 2 : 1	4 : 4 : 1	16 : 4 : 1
P_2O_5 in leaves immersed in water	100	100	100	100	100	100	100
Same, in leaves immersed in solutions	64	68	85	100	122	129	173

In the one of the experiments was traced the effect of additional doses of nitrogen fertilization of severed leaves upon the content of inorganic P_2O_5. The leaves were taken from plants which had been fertilized with nitrogen in different way; the petioles of the leaves were immersed into a weak solution of $NaNO_3$; the controls, in pure water. After several days the amount of inorganic P_2O_5 was determined in the leaves. It appeared that the leaves of the plants which had grown under insufficient nitrogenous nutrition showed under the influence of supplementary nitrogen fertilization a reduced content of inorganic P_2O_5 as compared with those which had been immersed into water; in the leaves taken from plants grown on a liberal nitrogen diet the reverse relations were observed.

Identical results have been obtained when supplementary doses of fertilizers were introduced by pulverizing the plants with the nutrient solutions.

The results here reported make one believe that from the analysis of young leaves for the content of inorganic P_2O_5 it is possible to form an idea of the type of nutrition of the plants.

It is an agreeable duty to thank Professor D. A. Sabinin for his valuable suggestions made in the course of this work.

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EXPERIMENTAL MORPHOLOGY

RESPIRATORY QUOTIENT IN REGENERATING TISSUES

By D. E. RYVKINA

(Communicated by I. I. Schmalhausen, Member of the Academy, 10. V. 1945)

In our previous communication (1) data have been presented concerning the decrease of the respiratory quotient in regenerating tissues of limbs of axolotl as studied by determining the content of free carbonic acid. In the tissues of normal limbs of 6—8-month-old axolotl the QR was found to be equal to 0.75, while the QR of a regenerate at the blastema stage was found to drop to 0.46 on the average. This decrease was compared with the low magnitudes of QR reported for the early development stage of frog eggs. According to Parnas, this reduction is due to the rearrangement of protein substances which takes place at these stages of development.

As not all carbonic acid produced by respiration is released as a free acid and the amount of the fixed carbonic acid was not determined in our preceding investigation, we still were facing the problem as to the total amount of carbonic acid produced by respirating tissues at different stages of regeneration, and the connexion between the metabolic changes and the amount of fixed carbonic acid produced in the regenerating organ.

For their QR estimations Warburg and Krebs make use of three manometers. In the first manometer the amount of the O_2 consumed is determined. Into the side space of the second and third manometers 0.3 cm³ of $3N$ H_2SO_4 is introduced. Into the principal space of the third manometer H_2SO_4 is introduced at the beginning of the experiment, in order to determine the initial amount of the fixed carbonic acid ($t = 0$) which was present there before. In the second manometer takes place the consumption of O_2 and the production of free carbonic acid, the amount of the latter being determined according to the following equation:

$$X_{CO_2} = \left(h'' - \frac{X_{O_2}}{K'_{O_2}}\right) K''_{CO_2}$$

60 minutes after the beginning of the experiment H_2SO_4 is also introduced into the principal space of the second manometer, and the amount of carbonic acid, B, fixed during this period, is determined: $B = HCO'_3 + CO''_3$. The difference between the second and the first reading is denoted as ΔB. A positive or negative value of ΔB shows an increase or a decrease in the amount of fixed carbonic acid, respectively. B is added to the amount of free carbonic acid produced ($K = CO_2 + H_2CO_3$), and the total amount ($\sum CO_2$) of the carbonic acid evolved by the tissue is thus obtained.

According to Runnström (2), the determination of the magnitude of B may serve for estimating the amount of the acid products formed in the tissues. A decrease in the value of B shows that in the tissues acids are formed stronger than the carbonic

acid, and that the latter is replaced by them in chemical compounds. Being stronger than the carbonic acid, HS reacts with $NaHCO_3$ according to the following equation: $HS + NaHCO_3 \rightarrow NaS + H_2CO_3$.

The value of B is reduced to 0, or becomes even negative, when the amount of acids produced is increased. After this level has been reached, free carbonic acid not connected with respiration is produced. According to Runnström, the value of B corresponds to the actual amount of the acids formed (*fixe Säure*).

Respiratory Quotient of Tissues at Early Stages of Regeneration

	Q_{O_2}	K	$B_{t=0}$	$B_{t=60}$	ΔB	$\Delta\Sigma CO_2$	QR	$\frac{\Delta B}{\Delta\Sigma CO_2}$
Tissue of the limb of normal axolotls (6—7-month-old)	−1.68	+0.70	2.13	3.05	0.92	1.62	0.96	0.57
	−1.58	+1.04	3.40	4.02	0.62	1.66	1.05	0.37
	−1.11	+0.50	3.91	4.48	0.57	1.07	0.96	0.53
	−1.46	+0.68	3.03	3.97	0.94	1.62	1.10	0.58
	−1.37	+0.82	2.64	2.12	0.48	1.30	0.95	0.37
	−1.10	+0.22	1.65	3.63	0.98	1.20	1.09	0.81
Average	−1.38	+0.66	2.79	3.54	0.75	1.41	1.02	0.54
Epithelization stage (8 days)	−1.13	+0.80	2.14	1.85	−0.29	0.51	0.45	−0.56
	−0.94	+0.70	1.73	1.70	−0.03	0.67	0.71	−0.04
	−0.98	+0.56	1.32	1.35	+0.03	0.59	0.60	+0.05
Average	−1.02	+0.69	1.73	1.63	−0.10	0.59	0.59	−0.18
Blastema (17—20 days)	−1.47	+0.45	0.49	1.17	0.68	1.13	0.76	0.60
	−1.00	+0.60	0.83	0.90	0.07	0.67	0.67	0.10
	−1.90	+1.06	0.64	0.70	0.06	1.12	0.58	0.05
	−1.86	+0.65	3.03	3.52	0.49	1.14	0.61	0.43
	1.76	+0.75	1.28	1.45	0.17	0.92	0.52	0.18
	−1.27	+0.23	1.68	1.85	0.17	0.40	0.31	0.42
Average	−1.54	+0.62	1.32	1.59	0.27	0.89	0.57	0.30
Stage of internal differentiation (26—29 days)	−2.24	+1.82	1.90	2.02	0.12	1.94	0.86	0.06
	−1.82	+0.75	1.59	2.28	0.69	1.44	0.78	0.41
	−2.00	+0.86	2.18	2.78	0.60	1.46	0.73	0.41
	−2.56	+1.22	0.98	1.06	0.08	1.30	0.50	0.06
Average	−2.15	+1.16	1.66	2.03	0.37	1.53	0.72	0.24

It will be seen from the table that the real value of QR for normal tissues of axolotl limbs is equal to 1.02 on the average. 8 days after the amputation, when epithelization takes place, QR in the tissues of the organ rudiment is equal to 0.59. A value of QR equal to 0.57 was found after 17—20 days at the blastema stage. At the stage of internal differentiation an increase in QR takes place (QR=0.71).

Thus, the decrease in QR at early stages of regeneration, found in our earlier experiments by measuring the free CO_2, is likewise observed when comparising the absolute magnitudes of QR. The figures given in the table show that the decrease in the content of carbonic acid expelled in regenerating tissues is largely dependent on the decrease in ΔB. In normal tissues this magnitude is equal to 0.75. It becomes

negative at the stage of epithelization (—0.10), while at the stage of blastema and internal differentiation it is equal to 0.21 and 0.37, respectively.

Thus, in regenerating tissues stable acids are produced, displacing the fixed CO_2 from chemical compounds. A decrease in the $\Delta B/\Delta\sum CO_2$ relation also indicates that stable acids are produced in the tissues. In so far as the experiments were carried out in a pure oxygen atmosphere, production of acids in aerobic conditions seems to take place.

As may be seen from the table, the consumption of O_2 and the release of free and total CO_2 exceed the normal level at the stage of internal differentiation. At this stage the values of ΔB, $\Delta B/\Delta\sum CO_2$ and QR are somewhat increased as compared to preceding stages, but still remain below the normal level.

The differentiation of the tissues is thus connected with an increased intensity of respiration as compared to that of the normal tissues. However, the processes of breakdown continue at this stage also. Thus, the investigation of the respiratory quotient and fixed carbonic acid during regeneration confirms the importance of the breakdown processes in the development and differentiation of tissues. The strongest shift of metabolism towards the predominance of aerobic breakdown takes place at the early stages of development and tissue differentiation.

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ZOOLOGY

THE AGE OF WHALES AS DETERMINED FROM THEIR BALEEN APPARATUS

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Two methods for determining the age in whales have been worked out by the expedition on board the ship «Discovery»: from the ankylosis between the epiphyses and the centre of the vertebra, and another, according to the number of *corpora lutea* in the ovaries of the females. The former permits of an age estimate in sex-immature animals, the latter until the climacteric only (in finwhales up to the age of 20 years, in blue whales up to the age of 30 years only).

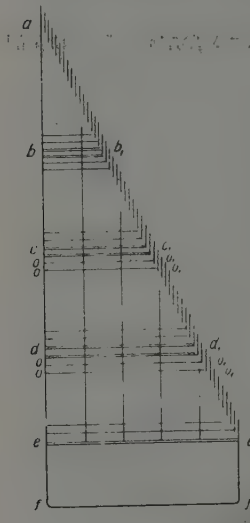

Fig. 1. Schematic drawing of a baleen plate of a four-year-old whale: abb_1, «smooth apex»; bb_1cc_1, cc_1dd_1, dd_1ee_1, «annual waves»: ee_1ff_1, part of baleen plate covered by the gum: oo_1, small transverse waves: $ab_1c_1d_1e_1$, fringe

In 1939 the present author suggested a method for a provisory age evaluation in whales according to the relief of the surface of their baleen plates (5). Somewhat later Ruud (3) has demonstrated the existence of a definite periodicity in the appearance of this relief. It has been suggested as far back as by Scorsby (4) and Eschricht and Reinhardt (1) that this relief was due to the seasonal conditions in their feeding.

Our investigation of baleen plates was carried on both at a whaling station and on museum collections; as a result we have found the two surfaces of baleen plates (the anterior, convex one; and the posterior, concave) of the various whale species to show the typical patterns described below. Investigated were *Balaena mysticetus* L., *Eubalaena glacialis glacialis* Bonn., *E. glacialis sieboldii* Gray, *Balaenoptera musculus* L., *B. physalus* L., *B. borealis* Less., *B. acutorostrata* Lacep., *Rhachianectes glaucus* Cope. The relief may have one of the following patterns: 1) longitudinal relief striation, 2) transverse relief undulation of two types, 3) «smooth apex», or distal part of the baleen plate, where the undulation is scarcely noticeable if at all.

Longitudinal striation is due to traces of concrescence of each smaller internal plate with a larger, external one along their contiguous fringe; it reflects the increment of the plate as a whole by breadth, and is of no use as a clue to the age of whales.

The transverse undulation consists of alternating small and big undulatory patterns (Fig. 1). The former consist of numerous impressions and protrusions, parallel to one another and transverse with respect to the longitudinal axis of the baleen plate. They may be of different relief. Their position with respect to the

external edge of the plate is not always perpendicular, being, as a rule, conformable to the line formed by the plane of the baleen plate on the spot of its immersion in the gum. The small undulations regularly become deeper at definite intervals, marking the boundaries between the annual increments of the plate and its base. If the baleen plate is viewed in the grazing light, these «annual segments» stand out against its surface as ridges and furrows, forming a pattern of rather gentle ripple marks. These are the big undulations for which we suggest the term of «annual» ones. The «annual wave» (or «step», as it is termed by Ruud) seems to include several smaller ones (15—30 in number), and may be separated from each adjoining wave according to its more pronounced fine ripple relief, or with the aid of grazing light. It is noteworthy that the «annual waves» are observable in each plate of the baleen apparatus, and are arranged at about the same distance from the gum, the level of their attachment having a gentle slope towards the anterior and posterior end of the set. When the apparatus is viewed from the side, they are observable as slanting convex lines, running along the external fringes of all plates of a given set, both right and left hand ones (Fig. 2, *bb*, *cc*, *dd*). Transverse undulation is the most pronounced near the base of the baleen plate, while towards the apex it is smoothed out, and may even vanish altogether; the same is also true of the longitudinal striation, though to a lesser degree. The area of such a «smooth» apex, as well as the degree of its smoothness, is subject to considerable individual variation.

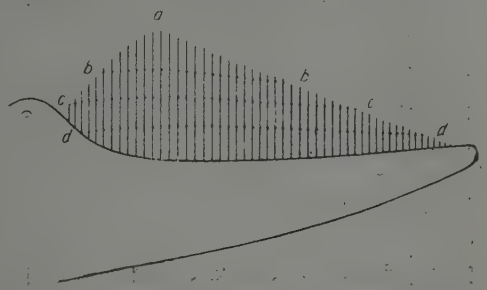

Fig. 2. Schematic drawing of baleen set of a four-year-old whale: *bb*, *cc*, *dd*, annual lines of baleen plate set

The formation of the undulatory pattern and of the «smooth apex» is interpreted by the present author in the following way. As soon as the young whale begins to feed by himself, his baleen plates undergo, first, continuous intense attrition of their apices and fringes, caused by the continuous friction against the tongue, food and water; secondly, the worne-out portions of the baleen are restored, sometimes in excess, owing to the continuous growth of the basal parts of the plates, the fringes being restored by way of a further splitting of the cortical and internal layer of the plate. In so far as the baleen undergoes attrition throughout the life of the animal, approximately the same quantity of matter being worn out yearly, the baleen plate does also grow continuously during the whole life of the animal, showing about the same increment by height within each annual cycle. The seasonal variations occurring in the feeding conditions within the annual cycle are reflected by the whole baleen apparatus in the form of transverse ripple marks of various intensity. The boundaries between the annual cycles of growth are therefore readily traceable when the baleen plate apparatus is viewed laterally. It should be pointed out that the surface of the baleen plates of sucklings and of young whales which have not yet attained sex maturity display the least pronounced relief and pattern*, which is observable for the «smooth apex» of adult whales only**. In young specimens this must evidently be due to 1) a higher rate of growth under uniform physiological conditions due to lactation, and 2) the continuous grinding by water suffered by the lateral surfaces of the baleen plate. Unlike this, in adult animals the «smooth

* Our collection contains a baleen plate of an almost mature foetus of finwhale (♂ 5.75 m long, obtained 18. X.1934, No. 322), bearing no traces of transverse undulation, and a baleen of a suckling finwhale at the close of the period of lactation (♀, 13.5 m long, obtained 6. VII. 1934, No. 118), showing a scarcely noticeable undulation.

** As may be seen from the curves given by Mackintosh and Wheeler [2] there is an abrupt jump in the rate of growth of whales after lactation.

Baleen Plates of Various Whale Species Investigated

Species	Kept at	Place of occurrence	Length of plate without fringes in cm	Thickness of plate in mm		Annual waves		Length of smooth apex in cm	Minimum age in years	Yearly increment of baleen in cm	Notes
				near gum	apex	number	interval in cm				
Balaena mysticetus	Polytechnical Museum	Arctics	350	9.5	4.0	18	15—17	50	19	15—20	
	Zoological museum of the Moscow University	Dejnev Cape	282	8.0	2,5	7	19—20	140	8		
	Same	Arctics	345	7.5	2.16	13	17—19	106	14		
	»	»	350	8.0	3.0	15	15—17	85	16		
Eubalaena glacialis glacialis	Anatomical Museum of the Moscow University	Europe seas	183	6.5	2.8	9	14—16	50	10	14—15	
Eubalaena glacialis sieboldii	The Darwin Museum	Far East	256	5.0	1.07	13	17—18	32	14	17—18	
Balaenoptera musculus	Polytechnical Museum, No. 260	Far East	75	5.5	3.5	5	12—13	11	6	10—13.5	Length of the whale 23.4 m
	Same	»	81	7.5	2.8	6	10—11	18	7		
	Moscow University, Ognev	Kamchatka	75	4.18	2.32	5	13.5	8	6		
Balaenoptera acutorostrata	The Darwin Museum	Europe	21	1.86	1.02	3	4.5	6	4	3.5—4.5	Length of the whale 6.2 m
	Polytechnical Museum No. 186	?	24	2.46	1.07	4	3—4	4	5		
Rhachianectes glaucus	Moscow University, Ognev	Chukotka peninsula	36	6.2	4.0	2	10	16	3	10	Length of the whale 12 m

apex» can represent nothing else than a remain of the plate surviving since the time of lactation, but somewhat shifted upwards and ground by the water, or else—if it has been entirely worne out—the smooth apex may be a mere result of the work of water. In the latter case the transverse undulation of the apex is usually much more pnonounced than in the former case. The accumulation of the «annual waves» proceeds till a definite age only, and is limited by a continuous wearing out of the plate's apex; this is why the «smooth apex» becomes shorter, as the animal advances in age.

We succeeded in determining the age of whales according to their baleen plates collected by us (see table) within the following age range: *Balaena mysticetus*, up to 19 years; *Eubalaena glacialis glacialis*, up to 10 years; *E. glacialis sieboldii*, up to 14 years; *Balaenoptera musculus*, up to 7 years; *B. acutorostrata*, up to 5 years; and *Rhachianectes glaucus*, up to 3 years.

As it is well known that the baleen plates of whales attain rather considerable dimensions by far exceeding those recorded in the table, it may well be assumed that age determinations based on their length could be practised also somewhat beyond the age range mentioned above. As shown by Ruud [3], the ratio of the medullary layer to the cortical layer changes with age to the advantage of the latter, as the plate growth thicker as a whole. In the case where the «smooth apex» is strongly worne out, certain corrections can therefore be introduced for very young specimens, based upon the thickness of the apex. In this way it is possible to form an idea of the dimensions of the portion of the baleen which has already been worne out by attrition, and of the number of annual ripple marks by which it had been ornamented. The highest rate of growth has been recorded for *Balaena mysticetus* (15—20 cm yearly); the lowest, for *Balaenoptera acutorostrata* (3.5—4.5 cm yearly).

From the data discussed above the following conclusions may be advanced, which, however, stand in need of checking by some other method of age evaluation.

1. The baleen plates of whales grow at the base and are subject to attrition both at the internal edge and the apex throughout the life of the animal.

2. In sex mature specimens the baleen plates show an annual accrescence by approximately the same magnitude, which may possibly be occasioned by similar conditions of annual wearing out of the plates.

3. The recurrent «steps» (or «annual waves») observable on the surface of the baleen plates reflect the seasonal course of feeding which is regularly reproduced from year to year; to each «step» corresponds an age interval of one year.

4. The baleen plates may only conventionally serve as a clue for an approximate estimation of the age and, because of their attrition, within a definite age range only.

5. For the purpose of age evaluation it is necessary to select the highest plates of the set, in which the accumulation of the «annual waves» is the best pronounced.

6. The lines which run along the external fringes of the baleen apparatus and are readily observed when the mouth of the animal is kept open, can serve as the most reliable clue to the determination of the annual cycles.

7. The age measured in years should be calculated by adding a unity to the number of the full «steps» (or waves) lying in the exposed portion of the plate, *i. e.* above the level of the gum. This mode of calculation is based upon the fact that the «smooth apex» wich does not enter the number of steps, develops within approximately half-a-year, while the part of the plate covered by the gum makes up about a half of an «annual wave». In the case the gum is adjoined by an incomplete step (*i. e.* a part of an annual wave), the number of years should be added a corresponding part of a year.

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